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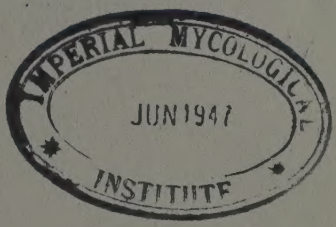

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TABLE DES MATIÈRES

Pages

MATHÉMATIQUES

G. Chogoshvili. On Duality Relations in Topological Spaces 131
Z. Shapiro. On Elliptical Systems of Partial Differential Equations 133

PHYSIQUE

D. I. Blokhintzev. Propagation of Sound in Turbulent Flow 136
B. M. Wul, Corresponding Member of the Academy, and I. M. Goldman. Dielectric Constant of Titanates of the Metals of the Second Group . . . 139

PHYSIQUE APPLIQUÉE

A. Andronow et N. Bautin. Stabilisation de route d'un avion neutre par autopilote ayant un servomoteur à vitesse constante et à zone d'insensibilité 143

CHIMIE

B. A. Arbuzov, Corresponding Member of the Academy, V. S. Abramov and L. A. Shapshinskaya. On the Reduction of Cyclones by Hydrocarbons . 147
N. D. Zelinsky, Member of the Academy, J. A. Arbuzov and M. I. Batuev. Contact Isomerization of Octene-1 and Investigation of Isomerization Products by the Combination Scattering Spectral Method 150

CHIMIE DU GLOBE

L. S. Berg, Corresponding Member of the Academy. On the Origin of the Ural Bauxites 154

GÉOLOGIE

S. G. Sarkissian. On the Age of the Ore-Bearing Strata of the Ulu-Telyak Manganese Deposit of Bashkiria (from Data of Mineralogical Analysis) . . 157

PHYSIOLOGIE VÉGÉTALE

N. N. Konstantinov. On the Influeuce of the Stock upon the Scion in Cotton 159

I. A. Krupenikov. Response of Certain Wild-Rose Species to High Salt Concentrations in Soils 162

A. A. Richter, Member of the Academy, **K. T. Sukhorukov** and **L.A. Ostapenko.** Photosynthesis and Growth 165

MORPHOLOGIE ÉVOLUTIONNAIRE

A. Rumjantzew and **L. Berezkina.** Further Observations on Transformations of Osteoclasts *in vitro*. Cultivation of Bone Tissue of Rat after Administration of Parathohormone 168

HELMINTHOLOGIE

T. S. Skarbilovich. Seasonal Destrobilization in the Cestode *Mesocestoides lineatus*, Parasitic of Silver-Black Fox 171

Édition de l'Académie des Sciences de l'URSS
Moscou 1945

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVI, № 4

MATHEMATICS

ON DUALITY RELATIONS IN TOPOLOGICAL SPACES

By G. CHOGOSHVILI

(Communicated by A. N. Kolmogoroff, Member of the Academy, 11. V. 1944)

1. Let R be a normal space, and M, a subspace of R. When M is open or closed in R, the r-dimensional (inner) homology ∇-group $\nabla^r(M)$ of M is defined * as the limit group of the direct spectrum $\{\nabla^r(K_\alpha - C_\alpha), \pi^\alpha_\beta\}$, where K_α is the nerve of a finite open covering O_α of the subspace M, C_α is the singular subcomplex of K_α, and $\nabla^r(K_\alpha - C_\alpha)$ denotes the r-dimensional Betti ∇-group $B^r_\nabla(K_\alpha - C_\alpha)$ with a discrete field of coefficients.

First of all, we give a definition of the two groups of M for the case where M is an arbitrary subspace of R.

Consider the totality $\{F_a\}$ of all closed sets F_a in R contained in M. We put $\nabla^r(F_a) < \nabla^r(F_b)$ when $F_a \subset F_b$, and the system $\{\nabla^r(F_a)\}$ becomes thus a non-bounded partially ordered set. In case $\nabla^r(F_a) < \nabla^r(F_b)$, we define a homomorphic mapping p^b_a of the group $\nabla^r(F_b)$ into the group $\nabla^r(F_a)$ as follows. We take an arbitrary bundle $h_b = \{h_{b\beta}\}$ of the group $\nabla^r(F_b)$, an arbitrary element $h_{b\beta}$ of this bundle and a cycle $z_{b\beta} \subset Z^r_\nabla(K_{b\beta} - C_{b\beta})$ from the class $h_{b\beta}$. The covering $O_{b\beta}$ of the set F_b whose nerve is $K_{b\beta}$ induces, in the sense of (1) 5, a covering $O_{a\alpha} = F_a \cdot O_{b\beta}$ of the set F_a. Consider $K_{a\alpha}$ as a subcomplex of the complex $K_{b\beta}$, and take a chain $z_{a\alpha}$ of the complex $K_{a\alpha}$ equal to $z_{b\beta}$ on $K_{a\alpha}$. We shall have $z_{a\alpha} \subset Z^r_\nabla(K_{a\alpha} - C_{a\alpha})$. Let $h_{a\alpha}$ be the homology class from $\nabla^r(K_{a\alpha} - C_{a\alpha})$ containing the cycle $z_{a\alpha}$, and h_a the bundle of the spectrum $\{\nabla^r(K_{a\alpha} - C_{a\alpha}), \pi^{a\alpha}_{a\beta}\}$ containing $h_{a\alpha}$. This definition of h_a depends neither on the choice of the cycle $z_{b\beta}$ in $h_{b\beta}$ nor of that of $h_{b\beta}$ in $h_b = \{h_{b\beta}\}$. We may, consequently, bring the bundle h_a in correspondence with the bundle h_b, and the homomorphism p^b_a is thus settled. These homomorphisms satisfy the condition: $p^b_a p^c_b = p^c_a$, so that the groups $\nabla^r(F_a)$ together with the homomorphisms p^b_a form an inverse spectrum $\{\nabla^r(F_a), p^b_a\}$. We now define the limit group $\nabla^r_F(M)$ of this spectrum, as usual (see (1), § 3), but introduce in $\nabla^r_F(M)$ the discrete topology (by a neighbourhood of a thread $\{h_a\}$ we mean a set of the threads $\{h'_a\}$ satisfying the condition: $h'_a \subset U_a$, where U_a are fixed neighbourhoods of the coordinates h_a of the thread $\{h_a\}$). The group $\nabla^r_F(M)$ will be called the r-dimensional closed homology ∇-group of the subspace M in the space R. If R is bicompact, then $\nabla^r_F(M)$ is a topological invariant of M. If M is closed in R, then $\nabla^r_F(M)$ is isomorphic to $\nabla^r(M)$.

2. Let $\{G_a\}$ be the totality of all open (in R) sets containing M. By introducing the relation of order in $\{\nabla^r(G_a)\}$: $\nabla^r(G_a) < \nabla^r(G_b)$, if $G_a \supset G_b$, we convert $\{\nabla^r(G_a)\}$ into a non-bounded partially ordered set. When $\nabla^r(G_a) < \nabla^r(G_b)$, we define a homomorphism q^b_a of the group $\nabla^r(G_b)$ into the group $\nabla^r(G_a)$ as

* For the notions and notations here used the reader is referred to (1).

follows. We take an arbitrary bundle $h_b=\{h_{b\beta}\}$ from $\nabla^r(G_b)$, an element $h_{b\beta}$ in h_b and a cycle $z_{b\beta}$ of the class $h_{b\beta}$: $z_{b\beta}\subset Z_\nabla^r(K_{b\beta}-C_{b\beta})$. Suppose a covering $O_{b\beta}$ of the open set G_b (the nerve of which is $K_{b\beta}$) to be induced, in the sense of ([2]), § 8, by a covering $O_{a\alpha}$ of the set G_a: $O_{b\beta}=G_b\cdot O_{a\alpha}$. Consider $K_{b\beta}$ as a subcomplex of the nerve $K_{a\alpha}$ of the covering $O_{a\alpha}$, and take a chain $z_{a\alpha}$ of the complex $K_{a\alpha}$ equal to $z_{b\beta}$ on the simplices $K_{a\alpha}$ belonging to $K_{b\beta}$ and equal to zero elsewhere. $z_{a\alpha}$ is a ∇-cycle on $K_{a\alpha}$ lying on $K_{a\alpha}-C_{a\alpha}$. Suppose that $h_{a\alpha}$ is the homology class from $\nabla^r(K_{a\alpha}-C_{a\alpha})$ containing $z_{a\alpha}$, and h_a is the bundle $\nabla^r(G_a)$ which contains $h_{a\alpha}$. The bundle h_a thus determined by $h_b=\{h_{b\beta}\}$ depends neither on the choice of the covering $O_{a\alpha}$ that induces the covering $O_{b\beta}$ on G_b, nor on the choice of $z_{b\beta}$ and $h_{b\beta}$ in $h_{b\beta}$ and h_b, respectively. Thus, h_a is brought in correspondence with h_b. Since this correspondence preserves the addition of bundles, we obtain a homomorphism q_a^b of the group $\nabla^r(G_b)$ into the group $\nabla^r(G_a)$. If $\nabla^r(G_a)<\nabla^r(G_b)<\nabla^r(G_c)$, then $q_a^bq_b^ch_c=q_a^ch_c$ for $h_c\subset\nabla^r(G_c)$. We thus obtain an inverse spectrum $\{\nabla^r(G_a),\ q_a^b\}$ whose limit group $\nabla_G^r(M)$ is called the r-dimensional open homology ∇-group of the set M in R. If M is an open set of R, then $\nabla_G^r(M)$ is isomorphic to $\nabla^r(M)$.

3. Let R be a normal locally-bicompact space simply connected with respect to the dimensions r and $r+1$ (see ([1]), § 7), M be an arbitrary subset of R. If $\{F_a\}$ is the system of closed (in R) sets contained in M, then $\{G_a\}$, where $G_a=R-F_a$, is the system of open sets in R containing $R-M$. Consider the spectra $\{\nabla^r(F_a),\ p_a^b\}$, and $\{\nabla^{r+1}(G_a),\ q_a^b\}$. By Kolmogoroff's duality law (see ([1]), § 8), the groups $\nabla^r(F_a)$ and $\nabla^{r+1}(G_a)$ are isomorphic. Denote by s_a this isomorphism. The following relations can be proved: $s_ap_a^b=q_a^bs_b$, $s_a^{-1}q_a^b=p_a^bs_b^{-1}$.

Hence the limit groups of the above spectra are isomorphic. We thus see that the r-dimensional closed homology group of a set M is isomorphic to the $(r+1)$-dimensional open homology group of the complement $R-M$: $\nabla_F^r(M)\cong\nabla_G^{r+1}(R-M)$. This isomorphism, evidently, yields Kolmogoroff's duality law, when M is a closed set.

4. The homomorphisms p_a^b defined above are valid also in the case where the group mapped is the homology group of an open set. This enables to map the group $\nabla_G^r(M)$ homomorphically into the group $\nabla_F^r(M)$. In fact, let $h_G=\{h_{G_a}\}$ be an element of $\nabla_G^r(M)$, where $h_{G_a}\subset\nabla^r(G_a)$, and G_a are open sets containing M. For every closed set $F_l\subset M$ the homomorphism p_l^a of the group $\nabla^r(G_a)$ into $\nabla^r(F_l)$ determines an element $p_l^ah_{G_a}=h_{F_l}$ of the group $\nabla^r(F_l)$. The totality of such h_{F_l} chosen singly in every $\nabla^r(F_l)$ forms a thread $\{h_{F_l}\}=h_F$ of the group $\nabla_F^r(M)$. Thus, $\nabla^r(G_a)$ is mapped homomorphically into $\nabla_F^r(M)$ for every a. Any other coordinate h_{G_b} of the thread h_G is mapped into the same thread h_F. Consequently, by associating the thread h_F to the thread h_G we obtain the required homomorphism of the group $\nabla_G^r(M)$ into $\nabla_F^r(M)$.

The isomorphism $\nabla_F^r(M)\cong\nabla_G^{r+1}(R-M)$ and the homomorphism of $\nabla_F^{r+1}(R-M)$ into $\nabla_F^{r+1}(R-M)$ generate a homomorphic mapping t of the group $\nabla_F^r(M)$ into the group $\nabla_F^{r+1}(R-M)$. If $\nabla_{F0}^r(M)$ is the kernel of the homomorphism t, $\cdot\nabla_F^r(M)$ is the quotient group of the group $\nabla_F^r(M)$ by $\nabla_{F0}^r(M)$, and $\nabla_F^{r1}(R-M)$ is the image of $\nabla_F^r(M)$ in the group $\nabla_F^{r+1}(R-M)$ under t, then we have the isomorphism $\cdot\nabla_F^r(M)\cong\nabla_F^{r1}(R-M)$. The groups $\cdot\nabla_F^r(M)$ and $\nabla_{\cdot F}^r(M)$ can also be defined directly, independently of the homomorphism t.

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MATHEMATICS

ON ELLIPTICAL SYSTEMS OF PARTIAL DIFFERENTIAL EQUATIONS

By Z. SHAPIRO

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We consider systems of partial differential equations of the form

$$\sum_{j=1}^{2n}\sum_{k=1}^{3} a_{ij}^{k}\frac{\partial u^{j}}{\partial x_{k}}=0 \qquad (1)$$

or in the form of matrices

$$A_1\frac{\partial \vec{U}}{\partial x_1}+A_2\frac{\partial \vec{U}}{\partial x_2}+A_3\frac{\partial \vec{U}}{\partial x_3}=0 \qquad (1')$$

where A_k is a constant matrix of the order $2n$, and $\vec{U}(x_1, x_2, x_3)$ is a $2n$-dimensional vector.

Such a system is called elliptic, if the form of order $2n$ with respect to the variables s^1, s^2, s^3 given by the determinant

$$Q(s^1, s^2, s^3)=|s^1A_1+s^2A_2+s^3A_3| \qquad (2)$$

is positive definite.

In the present note we consider the boundary problem for an elliptical system of the form (1) and construct the fundamental solutions of this system in the form of Abelian integrals.

Denote the components of the solution vector $\vec{U}(x_1, x_2, x_3)$ by $u^1, u^2, \ldots, u^{2n}$.

There exists a solution of the system (1) *defined in the half-space* $x_3>0$ *and such that for* $x_3\to 0$ n *arbitrary independent linear forms in* $u^1, u^2, \ldots, u^{2n}$ *tend to arbitrary continuous functions* $f_1, f_2, \ldots, f_{2n}$ *defined in the plane* (x_1, x_2) *and bounded on infinity.*

1. Consider the curve $Q(s^1, s^2, s^3)=0$ and a point with coordinates s_0^1, s_0^2, s_0 on this curve. If $\vec{U}_0(s_0^1, s_0^2, s_0^3)$ is a vector which is annuled by the matrix $\|s_0^1A_1+s_0^2A_2+s_0^3A_3\|$, then the vector $\Phi\left(\sum_{k=1}^{3} s_0^k x_k\right)\vec{U}_0(s_0^1, s_0^2, s_0^3)$, the scalar differentiable function $\Phi(t)$ being arbitrary, represents an elementary solution of the system (1). Put $\Phi(t)=1/t^2$. Then to every real value s^1, s^2 there correspond $2n$ complex pairwise conjugated elementary solutions

$$\vec{U}_1, \vec{U}_2, \ldots, \vec{U}_n, \quad \bar{\vec{U}}_1, \bar{\vec{U}}_2, \ldots, \bar{\vec{U}}_n \qquad (3)$$

We may without restraining the generality at first assume that the linear forms, whose values are given in the plane (x_1, x_2), coincide with the first n functions $u^1, u^2, \ldots, u^n$, and that $f_k \neq 0$ only for $k=1$.

Generally speaking *, for every s^1/s^2 it is possible to choose a real linear combination of the vectors (3), for which $u^1 \to \frac{1}{s^1 x_1 + s^2 x_2}$, $u^2, u^3, \ldots, u^n \to 0$, for $x_3 \to 0$.

The integral of this combination over all real values of the ratio s^2/s^1 is an Abelian integral over a closed cycle Γ on the Riemann surface of the integrated function. This integral we shall call the fundamental solution of the boundary problem stated above.

Denoting s^2/s^1 by s, and s^3/s^1 by t, and the coefficients of the chosen real combination by $c_\nu(s)$, we have the following formula for the fundamental solutions:

$$\vec{U}_\Phi(x_1, x_2, x_3) = \int_{-\infty}^{\infty} \sum_{\nu=1}^{n} \left\{ \frac{c_\nu(s)}{(x_1 + s x_2 + t_\nu x_3)^2} \vec{U}_\nu(s) + \frac{\bar{c}_\nu(s)}{(x_1 + s x_2 + \bar{t}_\nu x_3)^2} \bar{\vec{U}}_\nu(s) \right\} ds \qquad (4)$$

2. Replacing the cycle Γ by a homologic combination C of cycles γ_e of an adequately chosen basis of homologies $\left(C = \sum_e \alpha_e \gamma_e \sim \Gamma\right)$, and the integral over Γ by the integral over C and the sum of residues of the integrated function in the poles situated in the domain with the boundary $C - \Gamma$, we establish the following properties of the fundamental solution for $x_3 \to 0$.

Put $x_1 = x - \xi$, $x_2 = y - \eta$, $x_3 = z$, $R^2 = (x-\xi)^2 + (y-\eta)^2 + z^2$.

I. The first n functions of the fundamental solution

$$u^j_\Phi(x-\xi, y-\eta, z)$$

are representable in the form

$$\frac{z}{R^2} H^j\left(\frac{x-\xi}{R}, \frac{y-\eta}{R}, \frac{z}{R}\right) \qquad (j = 1, 2, \ldots, n)$$

where the H^j are continuous and bounded functions of their arguments for $z/R \geqslant 0$. Hence follows the absolute integrability of $u^j_\Phi(x-\xi, y-\eta, z)$ with respect to ξ and η for $z \neq 0$.

II. Put

$$J_\mu(z) = \int_{-\infty}^{\infty} \int_{-\infty}^{\infty} u^\mu_\Phi(x-\xi, y-\eta, z)\, d\xi\, d\eta \qquad (\mu = 1, 2, \ldots n)$$

$$J^G_\mu(z) = \iint_G u^\mu_\Phi(x-\xi, y-\eta, z)\, d\xi\, d\eta \qquad (\mu = 1, 2, \ldots n)$$

We easily see that

$$J_\mu(z) = \text{const} = J_\mu = \begin{cases} 4\pi^2 & \text{for } \mu = 1 \\ 0 & \text{for } \mu = 2, 3, \ldots n \end{cases}$$

$$\lim_{z \to 0} J^{\Delta_0}_\mu(z) = J_\mu,$$

where Δ_0 is defined by the inequalities $\begin{cases} |x-\xi| \leqslant a \\ |y-\eta| \leqslant b \end{cases}$

$$\lim_{z \to 0} J^{\Delta}_\mu(z) = 0$$

where Δ is either the complement of the domain of the form Δ_0 to the whole plane (ξ, η), or a finite domain not containing the point (x, y).

* If this does not hold for the plane $x_3 = 0$, it holds for arbitrarily near planes. Solving the problem for near domains by means of a linear transformation of the independent variables and passing to the limit, we obtain the solution for the half-space $x_3 > 0$.

3. Put $\frac{1}{4\pi^2} u^j_\Phi = R_{j1}$ (the index 1 indicates the specific character of the boundary values) and consider the matrix

$$R_{j\nu} \qquad (j=1, 2, \ldots, n;\ \nu=1, 2, \ldots, n)$$

Put now for $j \leqslant n$:

$$u^j(x, y, z) = \sum_{\nu=1}^{n} \int_{-\infty}^{\infty} \int_{-\infty}^{\infty} R_{j\nu}(x-\xi,\ y-\eta,\ z)\, f_\nu(\xi, \eta)\, d\xi\, d\eta \tag{5}$$

or $n < j \leqslant 2n$:

$$\frac{\partial u^j(x, y, z)}{\partial x} = \sum_{\nu=1}^{n} \int_{-\infty}^{\infty} \int_{-\infty}^{\infty} \frac{\partial R_{j\nu}(x-\xi,\ y-\eta,\ z)}{\partial x} f_\nu(\xi, \eta)\, d\xi\, d\eta \tag{6}$$

and similarly for the two other derivatives.

The property I of the fundamental solution implies the existence of the integrals (5). The existence of the integrals (6) for $z \neq 0$ is evident.

The property II implies that the boundary conditions

$$\lim_{z \to 0} u_\nu(x, y, z) = f_\nu(x, y) \qquad (\nu = 1, 2, \ldots n)$$

are satisfied.

Thus, the formulae (5) and (6) completely solve the stated boundary problem for the half-space.

4. Let now a bounded convex domain D with a smooth boundary S be given in the space of independent variables x, y, z. We put the problem: to find a vector $\vec{U}(x, y, z)$ giving the solution of the system (1) in D and such that on S the first n components are given functions.

Let P denote an inner point of the domain D; Q, its boundary point; $T(Q)$, the tangent plane to S at the point Q. Fixing Q, we may construct the matrix of fundamental solutions corresponding to the half-space with the boundary $T(Q)$ containing the domain D.

Denote the elements of this matrix by $R^{T(Q)}_{j\nu}(P, Q')$, where $Q' \subset T(Q)$, $P \subset D$. In particular, Q' may coincide with Q, and under this condition the fundamental solutions become functions of two points $P \subset D$, $Q \subset S$:

$$R^{T(Q)}_{j\nu}(P, Q)$$

We shall look for the solution of the stated problem in the form

$$u^j(P) = \sum_{\nu=1}^{n} \iint_S R^{T(Q)}_{j\nu}(P, Q)\, \mu_\nu(Q)\, d\sigma \qquad (j = 1, 2, \ldots, 2n) \tag{7}$$

It may be easily established that the functions $u^j(P)$ defined by the formula (7) tend, as P tends to the point $Q_0 \subset S$, to

$$\sum_{\nu=1}^{n} \iint_S R^{T(Q)}_{j\nu}(Q_0, Q)\, \mu_\nu(Q)\, d\sigma + \delta^\nu_j \mu_\nu(Q_0)$$

and, consequently, will be the solutions of the stated boundary problem if the $\mu_\nu(Q)$ are determined from the system of Fredholm equations

$$\sum_{\nu=1}^{n} \iint_S R^{T(Q)}_{\mu\nu}(Q_0, Q) \mu_\nu(Q)\, d\sigma + \delta^\nu_\mu \mu_\nu(Q) = f_\mu(Q_0) \qquad (\mu = 1, 2, \ldots n) \tag{8}$$

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PHYSICS

PROPAGATION OF SOUND IN TURBULENT FLOW

By D. I. BLOKHINTZEV

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Even if a sound is propagating in the atmosphere under most favourable conditions (slight breeze or calm) it still undergoes a strong damping which considerably exceeds the damping due to the atomic structure of the gas (viscosity, thermal conductivity, Kneser effect) ([1],[2]). It appears to be a generally recognized opinion that this strong damping is caused by the turbulent state of the atmosphere. But there is no clear idea as to the mechanism of the phenomenon, and no formulae to compute the damping coefficient. Lately Obukhov ([3]) calculated the scattering of sound by a turbulent flow, but the formula he obtained involved an indefinite function of the ratio of the turbulence scale l to wave-length λ. Moreover, the very notion of turbulence scale in the case of wind is rather vague.

Soloveichik ([4]) proceeds on very different considerations. He adopts for the sound the notion of turbulent viscosity which is intented to replace the true viscosity of the air in the equation for the propagation of a sound wave. This unsatisfactory state of affairs about the calculation of sound-wave damping in a turbulent atmosphere calls for a more thorough inquiry.

In the first place one must take into account that those pulsations of the flow's velocity whose dimensions l are so great in comparison with the sound wave-length λ that the propagation of sound through them can be studied by the methods of geometrical acoustics, fail to produce a scattering of sound.

They merely bring about a change in the form of the rays and accordingly a general vibration of the intensity of sound at the plan where the receiver is located. Therefore, the velocity of the turbulent flow should be resolved into two components which may be described as macrocomponent $\vec{v}$ and microcomponent $\vec{u}$, respectively:

$$\vec{v} = \int_{q<q_0} e^{i(\vec{q}\vec{x})} d\vec{U}(\vec{q}), \quad \vec{u} = \int_{q>q_0} e^{i(\vec{q}\vec{x})} d\vec{U}(\vec{q}) \tag{1}$$

$\vec{v}$ also includes the mean velocity of the flow $\vec{v}_0$. We have $q_0 = k/\mu$, where $k = 2\pi/\lambda$ is the wave number of the sound, and μ is an abstract number $\gg 1$. Following Obukhov ([3]), we shall consider the scattering of sound from a parallelepiped L^3 assuming that $L \gg \lambda$, $L \ll 2\pi/q_0$. Then as an approximation the velocity $\vec{v}$ can be regarded as constant in the volume.

If we pass to a system of coordinates moving with velocity $\vec{v}$, we shall find the frequency of sound f to have changed by a small quantity $f\frac{v}{c}$; the frequencies of turbulent pulsations in this system are $\nu = u(l)/l$, where l is the scale of pulsations, and $u(l)$, the velocity of pulsations of scale l.

In the case of isotropic and homogeneous turbulence the «2/3» law ([5],[6]) takes place, according to which $u^2 = \text{const}\, l^{2/3}$ (where const $\cong 1\ \text{cm}^{4/3}/\text{sec.}^2$), so that we have $\nu \cong \text{const}^{1/2} l^{-2/3} \ll f$ for all practically important f. By virtue of this circumstance when the sound moves across a turbulent flow it is the instantaneous picture of the turbulence, not its course in time, that are of importance. So the notion of turbulent viscosity is inadequate in the case of sound. The tensor of turbulent stresses, introduced by O. Reynolds, with what the notion of turbulent viscosity is connected, is obtained as the average of turbulent pulsations for a given average flow. The averaging is carried out on the assumption that every change in the average flow is slower than the chance pulsations of the velocity, produced by turbulence. But we have seen that just the reverse is true in the case of sound wave ($f \gg \nu$).

The action of a turbulent flow on a sound wave must be reduced to the scattering of sound, just as light is scattered, while crossing a turbid medium (in both cases chance changes of the velocity of vibrations take place). We are now going to estimate the magnitude of this scattering. The equation of propagation of sound in a moving medium runs as follows ([7],[8]):

$$\frac{1}{c^2}\frac{\partial^2\varphi}{\partial t^2} - \Delta\varphi + \frac{2}{c^2}\left(\vec{\Delta V}, \vec{\nabla}\frac{\partial\varphi}{\partial t}\right) - \int^t (\vec{\nabla}\varphi, \vec{\Delta V})\, dt = 0 \tag{2}$$

where $\vec{V} = \vec{c} + \vec{u}$ is velocity of flow, and φ is quasi-potential of sound wave. Proceeding to the system of coordinates in which $\vec{v} = 0$, one gets

$$\frac{1}{c^2}\frac{\partial^2\varphi}{\partial t^2} - \Delta\varphi = -\frac{2}{c^2}\left(\vec{u}, \vec{\nabla}\frac{\partial\varphi}{\partial t}\right) + \int^t (\vec{\nabla}\varphi, \vec{\Delta u})\, dt \tag{2'}$$

Discarding the right-hand side, we shall find the solution φ_0 which represents the primary wave

$$\varphi_0 = Ae^{i[\omega t - k(\vec{n}_1 \vec{x})]} \tag{3}$$

n_1 is unit vector in the direction of propagation of the primary wave within the limits of the selected parallelepiped L^3. The complete solution is $\varphi = \varphi_0 + \psi$, where ψ is scattered wave. For large distances R from the parallelepiped ψ has the form

$$\psi = \frac{B}{R}\, e^{i(\omega t - kR)} \tag{4}$$

The amplitude of the scattered wave B can be computed by the ordinary approximate methods (cf. ([3])). Through the introduction of the scattering vector

$$\vec{K} = k(\vec{n} - \vec{n}_1), \quad K = 2k \sin\frac{\theta}{2} \tag{5}$$

(θ is angle of scattering, n is direction of the scattered ray) we obtain

$$B = \frac{A}{4\pi c}\int_{L^3} (2\vec{u'}\, k^2 + \vec{\Delta u'}, \vec{n}_1)\, e^{i(\vec{K}, \vec{x'})}\, dv' \tag{6}$$

and the damping coefficient $\varkappa$ is expressed by the formula

$$\varkappa = \frac{\int \overline{|B|^2}\, d\Omega}{A^2 L^3} \tag{7}$$

where the dash over $|B|^2$ denotes averaging over turbulent pulsations, and integration over $d\Omega$ means integration over all scattering angles.

The averaging of the quantity $|B|^2$ leads to the expression

$$\overline{B^2} = \frac{A^2}{16\pi^2 c^2}\int\int dv'\, dv''\, e^{i(\vec{K}, \vec{\rho})}\, \overline{\{2u_1' k^2 + \Delta u_1'\}\{2u_1'' k^2 + \Delta u_1''\}} \tag{8}$$

where ρ is the distance between the points x' and x'', and u_1 is the projection of u on to the direction of the propagation of the sound wave n_1.

Substituting for $\vec{x}'$ and $\vec{x}''$ the relative coordinates $\vec{\rho}$ and the coordinates of the centre of gravity, we get:

$$\overline{|B|^2} = \frac{|A|^2 L^3}{16\pi^2 c^2} \int d v_\rho e^{i(\vec{K},\vec{\rho})} \{4k^4 M_{11}(\vec{\rho}) + 4k^2 \Delta M_{11}(\vec{\rho}) + \Delta^2 M_{11}(\vec{\rho})\} \quad (9)$$

where $M_{11}(\vec{\rho})$ is one of the moments of the correlation $M_{ik}(\vec{\rho}) = \overline{u_i' u_k''}$. $M_{ik}(\vec{\rho})$ can be represented through a spectral function

$$M_{ik}(\vec{\rho}) = \int\limits_{q>q_0} \psi_{ik}(\vec{q}) e^{i(\vec{q},\vec{\rho})} d v_q \quad (10)$$

For isotropic turbulence

$$\psi_{ik}(\vec{q}) = \left(\delta_{ik} - \frac{q_i q_k}{q^2}\right) f(q) \quad (11)$$

Hence for the energy of the microcomponent we have

$$\frac{1}{2}\overline{u^2} = \frac{1}{2} \operatorname{Sp} M_{ik}(0) = \frac{1}{2} \int\limits_{q>q_0} \operatorname{Sp} \psi_{ik}(\vec{q}) d v_q = 4\pi \int\limits_{q>q_0} f(q) q^2 dq \quad (12)$$

On the other hand, the relation for isotropic turbulence according to Obukhov ([5], [6]) is

$$\frac{1}{2}\overline{u^2} = \sqrt[3]{2}\left(\frac{D_0}{\varkappa}\right)^{2/3} q_0^{-2/3} \quad (13)$$

where D_0 is the energy dissipated in the turbulent flow under settled conditions ($\varkappa$ is a number of order 1). Comparing (12) with (13) gives

$$f(q) = \frac{2\sqrt[3]{2}}{12\pi}\left(\frac{D_0}{\varkappa}\right)^{2/3} q^{-11/3} = \gamma q^{-11/3} \quad (14)$$

From (7), (9), (10), (11) and (14) we obtain

$$\alpha = \frac{2\pi k^4}{c^2} \int \left(1 - \frac{K^2}{k^2} + \frac{K^4}{4k^4}\right)\left(1 - \frac{K_1^2}{K^2}\right) \gamma K^{-11/3} d\Omega \quad (15)$$

Integration over angles takes place in the domain $K > q_0$. Introducing the variable $\xi = \sin(\vartheta/2)$. $d\Omega = 4\xi\, d\xi\, d\varphi$, and noticing that integration over ξ in (15) extends from $\xi = 1/2\mu$ ($K = q_0$) to $\xi = 1$, we find

$$\alpha = \mu^{5/3}\beta\left(\frac{2\pi\gamma^{1/2}\lambda^{1/3}}{c}\right)^2 \frac{1}{\lambda};\ \beta = \frac{3}{5}(2\pi)^{1/3}\{1 + 25\,(2\mu)^{-1/3} - 21\,(2\mu)^{-4/3} + O\,(\mu^{-4})\} \quad (16)$$

Proceeding from $D_0 \cong 5\ \mathrm{cm^2/sec^3}$. (cf. ([6])), we get $2\pi\gamma^{1/2} = 3$. Measurements by Obukhov and Ershova for a slight wind give $2\pi\gamma^{1/2} \cong 6$. Still higher values have been obtained by Krassilnikov ([9]).

Turbulence does not seem to be perfectly isotropic and varies with the velocity of the wind. To draw a comparison with Sieg's experiments ([2]) we will take $2\pi\gamma^{1/2} \cong 7$. For $f = 500$ Hz ($\lambda = 68$ cm) we have, according to his measurements, $\alpha \cong 10^{-5}\ \mathrm{cm^{-1}}$. This gives a reasonable value for the parameter $\mu \cong 10$. As for the dependence on the wave-length, it will be expressed, according to (16), in the form $\lambda^{-1/3}$.

Sieg's measurements point to a similar dependence when the velocity of the wind is several meters per second. For a slight wind he failed to detect any dependence on the wave-length. But this stands in no contradiction to the results reported here, since the accuracy of his measurements was not high enough.

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PHYSICS

DIELECTRIC CONSTANT OF TITANATES OF THE METALS OF THE SECOND GROUP

By B. M. WUL, Corresponding Member of the Academy,

and I. M. GOLDMAN

In connexion with the study of a problem of practical interest, preliminary investigations of the dielectric properties of magnesium and calcium titanate have been carried out recently in our laboratory. It was shown that the dielectric constant of magnesium titanate ranges from 12 to 16, and that of calcium titanate is about 5 times greater (1). This sharp difference in the values of the dielectric constant might as well be explained by the high polarizability of the calcium ion compared to that of the magnesium ion, as by the fact that the lattices are of different type. The polarizability of the Mg^{++} ion is $0.10 \cdot 10^{-24}$ cm^3; that of the Ca^{++} ion is $0.47 \cdot 10^{-24}$ cm^3 (2).

Magnesium titanate forms a lattice of the ilmenite type; the lattice of calcium titanate is of the perovskite type.

In order to evaluate the part played by the change in composition and that played by the change in structure, we have obtained cadmium titanate which, like magnesium titanate, has the ilmenite type of lattice. Though the polarizability of Cd^{++} ion is more than two times as strong as that of Ca^{++} ion, yet we have found the dielectric constant of cadmium titanate only slightly to exceed 60.

It was thus established that the lattice of the perovskite type favours large dielectric constants.

In order to establish the general rule, titanates of strontium and of barium were prepared and investigated. They likewise form crystalline lattices of the perovskite type; titanates of beryllium and zinc which also belong to the second group, were likewise studied.

The data on the dielectric constant of all these compounds, as obtained by us, are brought together in Table 1. In it the titanates are separated into two groups according to the type of their crystalline lattice.

Table 1

Dielectric Constant of Titanates

Group	Metal in titanate	Lattice type	Dielectric constant
I	Beryllium	—	70
	Calcium	Perovskite	115
	Strontium	»	155
	Barium	»	$1150 < \varepsilon < 3200$
II	Magnesium	Ilmenite	17
	Zinc	—	30
	Cadmium	Ilmenite	62

The data on the crystalline structure of titanates have been borrowed from «Strukturbericht», a reference book. In the case of beryllium and zinc titanates the structure has not been established. According to their values of the dielectric constant, titanate of beryllium has been classed with the first group, and titanate of zinc with the second.

The values of the dielectric constant given in the Table 1 were measured at room temperature within a frequency range of 50 to $4 \cdot 10^6$ Hertz. Within this range the dielectric constant of all titanates, except the barium titanate, is independent of the frequency. In the case of barium titanate a slight increase of the dielectric constant is observed even at 5000 Hertz, and at a frequency of 50 Hertz the increase is rather conspicuous. This may be accounted for by additional polarization arising from electric conductivity. Yet for frequencies ranging from $74 \cdot 10^3$ to $4 \cdot 10^6$ Hertz—which was the interval used in our experiments—we have found the dielectric constant of barium titanate to be practically invariable with the frequency. To this interval of frequencies refers the lower value of ε given in Table 1.

This value relates to samples in which about 15 per cent of the volume were pores. But the dielectric constant of a sample increases with its density. So it may be taken for granted that the dielectric constant of barium titanate is above 1150.

It will be obvious from the above data that those titanates whose crystalline lattice is of the perovskite type have higher dielectric constants. The latter is especially high in the case of barium titanate. Among the dielectrics—except those from Rochelle salt, which exhibit a particular kind of polarization—the first place in the order of dielectric permeability is held by rutile. For rutile we have $\varepsilon = 173$ in the direction parallel to the main axis of the crystal, $\varepsilon = 89$ in the perpendicular direction. In randomly oriented crystals $\varepsilon = 114$. Barium titanate polycrystals have a dielectric constant exceeding 1150, *i. e.* 10 times greater than in the case of rutile polycrystals.

It may be seen from Table 1 that within either group of crystals the dielectric constant increases with increasing polarizability of the ions. Along with this in lattices of the perovskite type another factor comes into prominence—the increase of atomic or ionic polarizability due to the fact that the bond between the ions weakens as we pass from calcium titanate to barium titanate.

A lattice of the perovskite type (titanates of calcium, strontium and barium) has a cube for its elementary cell. The ion of the alkaline-earth metal is located in the centre of the cube, while the corners are occupied by the titanium ions, and in the middle of every edge there is an oxygen ion.

As the radius of the alkaline-earth metal ion grows, so does also the distance between the ions of titanium and oxygen, and the bond between these ions weakens. In the given sequence of titanates the ratio of the distance between the titanium ion and the oxygen ion to the sum of these ions radii can serve as a measure of rigidity if we adopt for the ionic radii fixed values according to Goldschmidt. The respective data are given in Table 2.

Table 2

Titanate of	Edge of elementary cube in Å	Distance between oxygen ion and titanium ion in Å	Sum of oxygen and titanium ion radii in Å
Calcium	3.80	1.90	1.96
Strontium	3.89	1.95	1.96
Barium	3.97	1.99	1.96

As is seen from Table 2, it is only for barium titanate that the ratio of the distance between the ion centres to the sum of their radii exceeds one. From this it may be inferred that the ionic bond is less rigid in barium titanate than in the other crystals, and, accordingly, the ionic polarizability is higher. This circum-

stance in conjunction with the stronger electronic polarizability of barium and oxygen ions accounts for the high value of the dielectric constant.

Undoubtedly, the predominant factor in the polarizability of barium titanate is the ionic polarizability of the crystalline lattice. Because of this, the behaviour of the polarizability with changing temperature is quite singular. The variation of the dielectric constant of barium titanate with temperature is brought out by Fig. 1. These data have been obtained for a frequency of $0.4 \cdot 10^6$ Hertz. Moreover, the frequency $1.5 \cdot 10^6$ Hertz has been used in the measurements carried out within a temperature interval of $+20$ to $+240°$ C, but the results were similar.

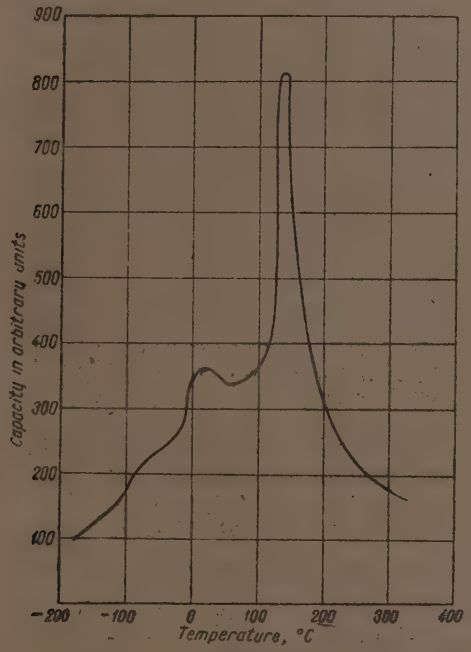

Fig. 1.

As the samples of barium titanate prepared by us were porous samples, an increase of capacity connected with the ionization of the air-filled interspaces was observed when the measurements were made at a high voltage. This is a familiar phenomenon in porous dielectrics in general. In the case of barium titanate it is expressed the more sharply, because the dielectric constant of this compound is exceptionally high. For one sample the change in capacity consequent upon an increase of voltage is shown in Fig. 2.

If the value of the dielectric permeability be computed from the capacity at high voltage, when the vacuous spaces in the sample are completely ionized, it will be found, proceeding from the geometrical dimensions of the sample, that $\varepsilon = 3200$. This value exceeds the magnitude of the dielectric constant in barium titanate, because the effective thickness of the condensor is less than its geometric thickness. Therefore, it may be taken for granted that at room temperature the dielectric constant of barium titanate is above 1150 and below 3200.

The high frequency measurements of the dielectric constant of barium titanate were carried on at a field strength of about $2-3\,V_{max}/cm$. As wide variations of voltage as 1.5 times did not practically affect the capacity. On the other hand,

measurements at the 50 Hertz frequency have shown that the capacity remains almost constant when the voltage is varied from 5000 to 21000 $V_{max/cm}$, after the ionization of air-gaps is complete.

Neither these facts nor the composition and structure of barium titanate permit it to be classed with the group of Rochelle salt dielectrics. The particular features of barium titanates are being investigated by the authors.

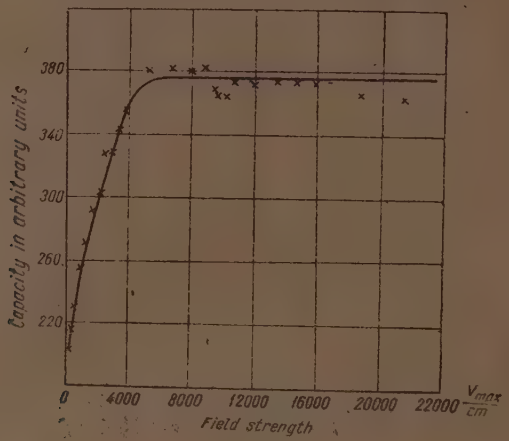

Fig. 2.

Measurements at high frequency on a Q-meter were carried out in an electroceramical plant with the participation of G. I. Skanavi. Control measurements at high frequencies and measurements at audio-frequencies were carried out in NIISKA with the participation of Kh. A. Valeyev.

The experimental work in the Institute was carried out with the assistance of junior scientific worker R. J. Razbash and laboratory assistant N. L. Stroganova.

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PHYSIQUE APPLIQUÉE

STABILISATION DE ROUTE D'UN AVION NEUTRE PAR AUTOPILOTE AYANT UN SERVOMOTEUR À VITESSE CONSTANTE ET À ZONE D'INSENSIBILITÉ

Par A. ANDRONOW et N. BAUTIN

(Présenté par L. I. Mandelstam, de l'Académie, le 31. V. 1944)

Dans une note précédente (1) nous avons étudié la théorie de la stabilisation de route d'un avion neutre (c'est-à-dire sans couple redresseur) par un autopilote muni d'un servomoteur à vitesse constante sans zone d'insensibilité. Le but de ce travail est de trouver les conditions d'amorçage et les caractéristiques quantitatives des auto-oscillations engendrées dans le même système par la présence d'une zone symétrique d'insensibilité du dispositif automatique *.

1. Désignons $2\psi_0$ ($\psi_0 > 0$) la largeur de la zone d'insensibilité. Reprenant les notations du travail précédent, on peut écrire les équations du mouvement du système étudié sous la forme:

$$\ddot{\varphi} + M\dot{\varphi} = -N\eta;\ \dot{\eta} = F;\ \psi = \varphi + \beta\dot{\varphi} - \frac{1}{a}\eta \qquad (1)$$

où $F = K$ pour $\psi > \psi_0$; $F = -K$ pour $\psi < -\psi_0$; $F = 0$ pour $-\psi_0 < \psi < \psi_0$ et $F = a(\dot{\varphi} + \beta\ddot{\varphi})$ pour $\psi = \psi_0$, $\dot{\psi} = -\psi_0$.

Faisant les changements de variables indiqués dans le travail précédent et supposant encore $A = M\beta > 0$, $B = \frac{M^2}{Na} > 0$, mettons le système (1) sous la forme:

$$\frac{dz}{dt'} = -z + f(u, z, \psi);\quad \frac{du}{dt'} = f(u, z, \psi);\quad \frac{d\psi}{dt'} = g(u, z, \psi) \qquad (2)$$

$$\begin{array}{llll} f \equiv -1, & g \equiv u + (A+1)z - B & \text{si} & \psi > \psi_0 \\ f \equiv 1, & g \equiv u + (A-1)z + B & \text{si} & \psi < -\psi_0 \\ f \equiv 0, & g \equiv 0 & \text{si} & |\psi| < \psi_0 \\ -Bf \equiv u + (A-1)z, & g \equiv 0 & \text{si} & \psi = \psi_0, \end{array}$$

$$0 < u + (A-1)z < B, \text{ et si } \psi = -\psi_0,\ -B < u + (A-1)z < 0.$$

Considérons u, z, ψ comme des coordonnées cartésiennes et étudions les solutions du système (2) déterminées par la condition de continuité de $u(t')$, $z(t')$, $\psi(t')$ aux points de discontinuité de $f(u, z, \psi)$ et $g(u, z, \psi)$. L'extension en phase u, z, ψ est remplie par des morceaux de trajectoires situés: 1) dans les plans $u = c = \text{const}$ pour $|\psi| < \psi_0$, 2) dans les surfaces cylindriques $z = ce^u - 1$ pour $\psi > \psi_0$, 3) dans les surfaces cylindriques $z = ce^{-u} + 1$ pour $\psi < -\psi_0$.

* Des problèmes similaires sont examinés dans les travaux de Kotelnikow (2), Boutenine (3), Boulgakov (4). La théorie des transformations ponctuelles des surfaces permet de résoudre le problème non-linéaire considéré sans supposer petite aucune des grandeurs figurant dans les équations du mouvement.

Les plans $\psi = \pm \psi_0$ sont les lieux de jointure des trajectoires. Sur les «lames» (σ): $\psi = \psi_0$, $0 < u + (A-1)z < B$ et (σ'): $\psi = -\psi_0$, $-B < u + (A-1)z < 0$, partagées elles-mêmes en trajectoires, s'opère la jonction de trajectoires de sens différents. Le point figuratif arrivé sur la lame poursuit son mouvement sur une des trajectoires de celle-ci, et ne la quitte qu'après en avoir atteint le bord. Sur les parties des plans n'appartenant pas aux lames se joignent des trajectoires de même sens. Le segment $u = z = 0$, $|\psi| \leqslant \psi_0$ est un segment de repos.

2. L'équation (2) détermine la transformation du plan $\psi = \psi_0$ (qui comprend la lame σ) en lui-même. Partageons le plan u, z $(\psi = \psi_0)$ en quatre morceaux:

$G^{(1)}$ — comprenant les points pour lesquels $u + (A-1)z - B > 0$ ou $u + (A-1)z - B = 0$, $u > A + B$;

$\overline{G}^{(2)}$ — comprenant ceux des points de la lame σ, qui en suivant les trajectoires, en lesquelles est divisée la lame, deviennent des points de l'arête (L): $u + (A-1)z = 0$, $u < 0$ $(\psi = \psi_0)$. L'arête L elle-même n'est pas comprise dans $G^{(2)}$.

$G^{(2)}$ — comprenant ceux des points de la lame σ, qui en suivant les trajectoires, en lesquelles est divisée la lame, deviennent les points de l'arête (M): $u + (A-1)z - B = 0$, $u > A + B$ $(\psi = \psi_0)$. L'arête M elle-même n'est pas comprise dans $G^{(2)}$.

$G^{(3)}$ — comprenant les points pour lesquels $u + (A-1)z < 0$ ou $u + (A-1)z = 0$, $z < 0$.

Appelons $H^{(1)}$, $\overline{H}^{(2)}, H^{(2)}$, $H^{(3)}$, les morceaux du plan $\psi = -\psi_0$ symétriques respectivement à $G^{(1)}$, $\overline{G}^{(2)}$, $G^{(2)}$, $G^{(3)}$ par rapport à l'origine des coordonnées, et L', M' les demi-droites symétriques à L et M.

Si le point initial $u = u_0$, $z = z_0$, $\psi = \psi_0$ $(\psi = -\psi_0)$ appartient à $\overline{G}^{(2)}(\overline{H}^{(2)})$, il parvient en vertu de (2) en un point de l'arête L (L') appartenant à $G^{(3)}(H^{(3)})$. Nous appelons cette transition du point u_0, z_0 *transformation* $E^+(E^-)$. Si le point initial appartient à $G^{(2)}(H^{(2)})$, il parvient en vertu de (2) en un point de l'arête M (M') appartenant à $G^{(1)}(H^{(1)})$. Nous appelons cette transition *transformation* $E^+(E^-)$. Si le point initial u_0, z_0 $(t' = 0)$ appartient à $G^{(1)}(H^{(1)})$ apres avoir quitté en vertu de (2) le plan $\psi = \psi_0$ $(\psi = -\psi_0)$ lorsque $t' = 0$, il regagne ce plan lorsque $t' = \tau$ au point u_1, z_1 appartenant à $G^{(2)}(H^{(2)})$, $\overline{G}^{(2)}(\overline{H}^{(2)})$ ou $G^{(3)}(H^{(3)})$. Nous appelons cette transition *transformation* $S^+(S^-)$. Si le point initial u_0, z_0, $\psi = \psi_0$ $(\psi = -\psi_0)$ appartient à $G^{(3)}(H^{(3)})$, il arrive lorsque $t' = \tau'$ au point u_1, z_1 du plan $\psi = -\psi_0$ $(\psi = \psi_0)$. Nous appelons cette transition *transformation* $S_0^-(S_0^-)$. Les transformations S^+ et S_0^+ s'expriment par les formules

$$(S^+) \quad \begin{cases} z_0 = (1+\zeta)e^{\tau} - 1; & z_1 = \zeta \\ u_0 = (A-1)(1+\zeta)\dfrac{1-e^{\tau}}{\tau} + \dfrac{\tau}{2} + A + B - 1; & u_1 = u_0 - \tau \end{cases} \tag{3}$$

$$(S_0^+) \quad \begin{cases} z_0 = \zeta; & z_1 = \zeta e^{-\tau'} \\ u_0 = \dfrac{(A-1)(e^{-\tau'} - 1)\zeta - 2\psi_0}{\tau'}; & u_1 = u_0 \end{cases} \tag{4}$$

3. S_0^+ transforme l'arête L en une courbe Γ située dans le plan $\psi = -\psi_0$.

$$u = \frac{2\psi_0}{1 - \tau' - e^{-\tau'}}, \quad z = \frac{2\psi_0}{(A-1)(1 - \tau' e^{\tau'} - e^{\tau'})} \quad (0 < \tau' < \infty) \tag{5}$$

Considérons la demi-droite $z = ku$, où k est défini par l'équation (6) si $4B-(A+B)^2>0$ ou (7) si $4B-(A+B)^2<0$:

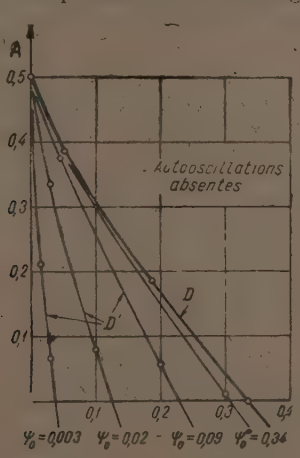
$$\frac{2(A+B)}{Be\sqrt{4B-(A+B)^2}}\operatorname{arctg}\frac{[(A-1)k+1]\sqrt{4B-(A+B)^2}}{(A-1)(A+B)k+A-B} = (A-1)[(A-1)k^2+(2-A-B)k-1] \quad (6)$$

$$\frac{2(A+B)}{Be\sqrt{(A-B)^2-4B}}\operatorname{arth}\frac{[(A-1)k+1]\sqrt{(A-B)^2-4B}}{(A-1)(A+B)k+A-B} = (A-1)[(A-1)k^2+(2-A-B)k-1] \quad (7)$$

Les points $\overline{H}^{(2)}$ sont divisés par la demi-droite $z=ku$ en trois classes: 1) les points appartenant à la demi-droite; la transformation $\overline{E}^-$ ne change pas leur distance au plan $u=0$; 2) les points pour lesquels $z<ku$; la transformation $\overline{E}^-$ les rapproche du plan $u=0$; 3) les points pour lesquels $z>ku$; la transformation $\overline{E}^-$ les éloigne du plan $u=0$.

Si $k>0$, la demi-droite $z=ku$ *coupe la courbe* Γ (l'intersection est unique). Supposons que l'intersection P appartient à $\overline{H}^{(2)}$ (ce qui est toujours le cas pour k ou ψ_0 suffisamment petits). Il résulte de la définition de k que le point P est invariant par rapport à la transformation $T^*=\overline{E}^-S_0^-\overline{E}^+S_0^+$.

Elle transforme tout point de $H^{(2)}$ choisi sur la courbe Γ en un point appartenant également à Γ et plus rapproché de P. La transformation T^* étant itérée, la suite des points de Γ converge vers P. T^* transforme



tout autre point de $\overline{H}^{(2)}$ en un point de la courbe Γ. Le point invariant est donc stable et le segment de repos instable. Le mouvement périodique stable correspondant est composé de morceaux de trajectoires de phase appartenant aux lames σ et σ' et de trajectoires comprises entre les plans $\psi=\psi_0$ et $\psi=-\psi_0$. Si $k<0$, la demi-droite $z=ku$ *ne coupe pas la courbe* Γ; la transformation T^* rapproche alors tout point de $\overline{H}^{(2)}$ du segment de repos. La transformation étant itérée, la suite des points converge vers l'extremité du segment de repos. Celui-ci dans ce cas est stable. Si $k=0$, la demi-droite $z=ku$ *est tangente* à la courbe Γ. Faisant dans (6) et (7) $k=0$ on trouve les valeurs de A et B pour lesquelles le segment de repos change la stabilité, et il s'en dégage un mouvement périodique stable (courbe D sur la figure).

4. Les caractéristiques quantitatives du mouvement périodique considéré peuvent être facilement calculées. La période $\tau^*=2(\tau_0+\tau_1)$, où τ_0 et τ_1

sont la durée du trajet entre les plans $\psi = \psi_0$ et $\psi = -\psi_0$ et celle du trajet sur les lames σ et σ', peut s'écrire sous la forme

$$\tau^* = \frac{2}{A+B} \ln \left\{ \frac{(A-1)[(A-1)k^2 + (2-A-B)k - 1]}{B[(1-A)k]^{A+B}} \right\}$$

k étant déterminé par (6) ou (7). τ^* croît indéfiniment lorsque le mouvement périodique converge vers le segment de repos ($k \to 0$).

5. L'intersection P de la demi-droite $z = ku$ et de la courbe Γ appartient au mouvement périodique (est invariante par rapport à $T^* = \bar{E}^- S_0^- \bar{E}^+ S_0^+$) tant qu'elle est dans $\bar{H}^{(2)}$. Si P sort de $\bar{H}^{(2)}$, le mouvement périodique comprend des morceaux de trajectoires faisant partie des surfaces cylindriques $z = ce^u - 1$ et $z = ce^{-u} + 1$. A ce mouvement périodique correspond sur le plan le point P' invariant par rapport à

$$T' = S^- \bar{E}^- S_0^- S^+ \bar{E}^+ S_0^+ \quad \text{ou} \quad T'' = E^- S^- \bar{E}^- S_0^- E^+ S^+ \bar{E}^+ S_0^+.$$

Les courbes D' (figure) calculées en partant de la condition que P se trouve sur la frontière $\bar{H}^{(2)}$, ψ_0 étant fixé, donnent sur le plan A, B la frontière entre les mouvements périodiques contenant et ne contenant pas de morceaux de trajectoires faisant partie des surfaces cylindriques.

6. Les résultats de ce travail concernant l'amorçage d'auto-oscillations dues à la zone d'insensibilité sont représentés par la figure.

Si $A = M\beta$ et $B = M^2/Na$ sont suffisamment grands, il n'y a pas d'auto-oscillations et il existe *un état d'équilibre indifférent [par rapport à ψ] à l'intérieur de la zone d'insensibilité.*

Si A et B diminuent de telle façon que le point représentatif sur le plan des paramètres A, B franchit la courbe D en se rapprochant de l'origine, les auto-oscillations apparaissent dans la zone d'insensibilité. *La courbe D est donc la frontière de la région où existent des auto-oscillations contenues à l'intérieur de la zone d'insensibilité.* Cette frontière ne dépend pas de ψ_0.

Fixons maintenant une valeur de ψ_0. Si A et B continuant à varier, le point représentatif sur le plan des paramètres A, B vient à franchir la courbe D' (correspondant à la valeur choisie de ψ_0), les auto-oscillations commencent à dépasser la zone d'insensibilité. *D' est donc la frontière de la région où existent des auto-oscillations dépassant la zone d'insensibilité.*

La courbe D et la famille des courbes D' montrent clairement le rôle de β en tant que facteur étouffant les auto-oscillations. Remarquons cependant qu'en choisissant a suffisamment petit on peut éviter les auto-oscillations même pour $\beta = 0$.

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CHEMISTRY

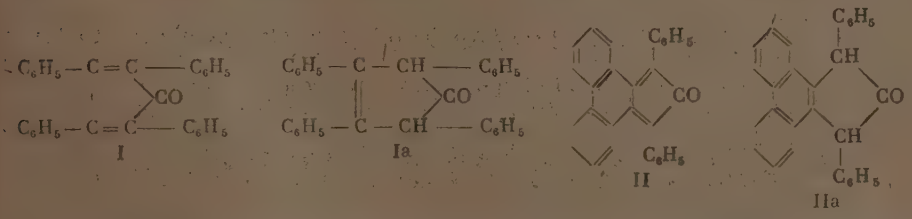
ON THE REDUCTION OF CYCLONES BY HYDROCARBONS

By B. A. ARBUSOV, Corresponding Member of the Academy,
V. S. ABRAMOV and L. A. SHAPSHINSKAYA

While studying the reactions of tetraaryl-cyclopentadienones — cyclones with various unsaturated compounds—the formation of colourless crystalline products has more than once been observed by the authors. The nature of these products does not depend on the unsaturated component, but solely on the cyclone used in the reaction. Special experiments have shown that these products can be obtained also by heating cyclones in inert solvents (xylene, toluene) even in the absence of unsaturated compounds.

The study of the nature of these colourless products has led to somewhat unexpected results: they have been found to be dihydro-derivatives of the corresponding cyclones.

Thus, from tetraphenyl-cyclopentadienone I dihydro-tetraphenyl-cyclopentadienone Ia was obtained. Phencyclone II yielded dihydrophencyclone IIa:



The structure of the dihydro-derivatives obtained has been established by means of mixed melting point determinations using synthetic products, prepared by reducing cyclones with hydroiodic acid according to Dilthey.

Experiments have shown that the formation of dihydrocyclones Ia and IIa takes place when cyclones I and II are heated in sealed tubes with benzene, xylene, toluene, tetraline and *n*-hexane.

Thus, it has been demonstrated that the hydrogenation of unsaturated compounds (cyclones) can be accomplished through dehydrogenation of saturated, hydroaromatic and aromatic compounds in the absence of hydrogenating and dehydrogenating catalysts.

The results of the experiment are given in the table.

As may be seen from the table, the disproportionation reaction between the hydrocarbons and cyclones proceeds at comparatively low temperatures, especially so in the case of the more reactive phencyclone. Cyclones appear to be strong acceptors of hydrogen at 200—230°, so much so in fact that even benzene can be used as a donor (phencyclone).

Solvent	Cyclone taken, g	Temperature of heating, °C	Time of heating, hrs	Yield of raw product of reaction		Dihydrocyclone recrystallized		Melting point of dihydrocyclone	
				g	%	g	%	raw	recrystallized
Phencyclone									
Xylene	0.5	260—280	1	0.1	20	—	—	288—290	316—318*
Toluene	2.0	300—330	4	1.6	80	—	—	295—300	316—318
Benzene	3.0	220—250	18	2.6	86	—	—	295—300	316—318
»	3.0	180—190 190—200 200—230	10 15 20	2.8	93	0.5	16.5	280—300	314—316**
Tetraline	1.0	147	16	0.55	55	0.2	20.0	280—300	315—317
n-hexane	0.5	180—210	16	0.4	80	0.25	50	280—300	313—315
Ethyl alcohol (absolute)	0.5	178—184	24	0.45	90	0.39	78	300—304	314—316
Tetraphenyl-cyclopentadienone									
Xylene	0.5	300	24	0.4	80	—	—	148—154	160—161***
Tetraline	0.5	140—150 180—200	20 8	0.35	70	0.15	30	143—148	161—162
Ethyl alcohol (absolute)	0.5	200—230 230—250	14 12	0.4	80			160—161	161—162

In literature one can find data on the hydrogenation-dehydrogenation reaction of hydroaromatic compounds in the absence of catalysts. Thus, abietic acid when heated to 260—270° partly undergoes disproportionation into dehydroabietic and hydroabietic acid (1). Interesting cases of dehydrogenation reactions at low temperatures of hydroaromatic compounds under the influence of chloranil and of other quinones having a high oxidative potential, have been described lately by Arnold and Collins (2).

The data quoted above point to the possibility of hydrogenation at the expense of hydrogen of the aromatic nucleus at a comparatively low temperature in the absence of catalysts. The possibility of hydrogenation of diaryl-acetylenes and diaryl-ethylenes by means of benzene at low temperatures in the presence of aluminium chloride has of late been described by Fuson et al. (5). The mechanism of the reaction remains obscure. Probably, it is a complex affair and does not consist in a direct addition of hydrogen. We have not cleared up as yet the character of the compounds formed in this reaction with cyclones from saturated, hydroaromatic and aromatic compounds. Work in this direction is being continued. The possibility of reducing cyclones by other hydrogen donors besides hydrocarbons without catalysers has been demonstrated using as an example the reduction of cyclones I and II with ethyl alcohol. Phencyclone is reduced to dihydrophencyclone IIa at 180°, and tetraphenyl-cyclopentadienone at 250°. Ethyl alcohol is dehydrogenated to acetaldehyde (the Schiff reaction with fuchsine, decolourised by sulphurous acid, 2,4-dinitrophenyl-hydrazone, m. p. 151—152°). The possibility of a conjugated dehydrogenation-hydrogenation reaction between alcohols and olefins in the presence of a catalyst (Ni) has been demonstrated by many authors (see, for instance(3)).

* Melting point of dihydrocyclone 316—318°. Mixed melting point 316—318°.

** Besides dihydrophencyclone a high melting product of unknown nature is formed.

*** Melting point of dihydro-tetraphenyl-cyclopentadienone 161—162°. Mixed melting point 161—162°.

The reaction between cyclones and hydrocarbons or ethyl alcohol was carried out by heating the mixture in sealed tubes. At the beginning the tubes were heated at low temperatures, and then if necessary the temperature was raised. The end of the reaction was determined from the disappearance of the cyclones' colour. In the case of phencyclone the dihydro-derivative usually separated as a crystalline powder. After the elimination of the solvent (in the case of tetraline through steam distillation) the crystalline powder was separated and recrystallized from toluene in the case of phencyclone II and from ethyl alcohol in the case of cyclone I. The melting point of dihydrophencyclone was determined using a copper bloc.

The synthesis of dihydrocyclones was carried out according to Dilthey by using hydroiodic acid and red phosphorus. The reduction of the cyclones with HI was achieved at the low temperature of 110—115° compared to 170°, as used by Dilthey. Reduction under lower temperatures gives better yields and a purer product.

Attempts to discover biphenyl in the case of benzene and naphtalene in the case of tetraline have been so far unsuccessful.

Conclusions. The hydrogenation of cyclones is possible by means of paraffinic, hydroaromatic and aromatic hydrocarbons, or by means of ethyl alcohol in the absence of catalysts at temperatures of 140—300°.

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CHEMISTRY

CONTACT ISOMERIZATION OF OCTENE-1 AND INVESTIGATION OF THE ISOMERIZATION PRODUCTS BY THE COMBINATION SCATTERING SPECTRAL METHOD

By N. D. ZELINSKY, Member of the Academy, J. A. ARBUZOV and M. I. BATUEV

It has been found previously ([1], [2]) that cyclohexene when brought into contact with certain oxides (alumina, beryllium oxide, silica gel, titanium dioxide) at 350—450°C undergoes isomerization into 1-methylcyclopentene-1*. Under the same conditions mono- and disubstituted homologs of cyclohexene undergo isomerization into cyclopentene homologs ([1], [3]). It has further been found ([4]) that cycloheptene when in contact with silica gel is isomerized into dimethylcyclopentene at 400°C.

It seemed interesting to find out whether olefins with a non-forked chain of carbon atoms would be isomerized under similar conditions into hydrocarbons of a branched structure. It was found ([2]) that heptene-1 when brought into contact with beryllium oxide at 450°C undergoes considerable isomerization into hydrocarbons of a branched structure. Hexene-1 in contact with silica gel at 400—450°C was also isomerized into hydrocarbons of a branched structure (unpublished data).

Thus, it has been shown that non-forked alkenes too were capable of isomerization into hydrocarbons of a branched structure if acted upon by oxides as contacts.

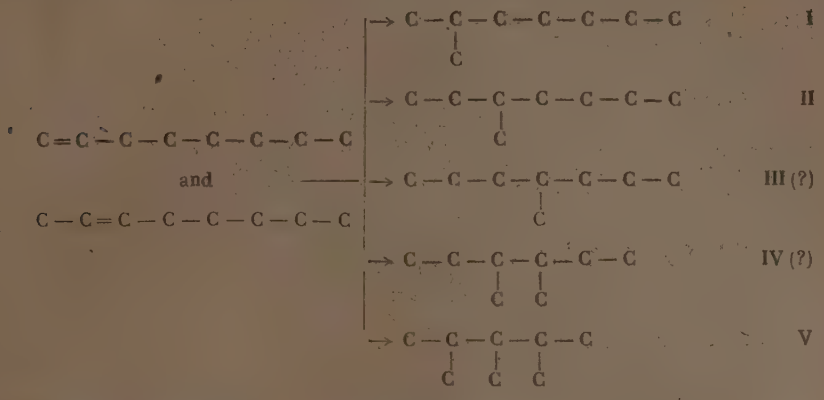
In the present paper we report some results of our experiments on the isomerization in the presence of oxides used as contacts of octene-1. It follows from the data reported that this olefin with a normal non-forked chain to a considerable extent undergoes isomerization into hydrocarbons with a branched chain of carbon atoms.

Petrov and Cheltzova ([5]) discovered in 1937 that *n*-octenes (mixture of isomers) in the presence of zinc chloride or phosphoric acid deposited on pumice at 300—400°C are isomerized into hydrocarbons of a branched structure.

In 1939 Egloff *et al.*([6]) studied the catalytic cracking of aliphatic hydrocarbons—among their number *n*-octenes (a mixture of octene-1 and octene-2)—in the presence of an activated silica-alumina mass (the composition and method of preparation of this contact are not described by the authors). It was found out that at 375 and 385°C the mixture of *n*-octenes underwent both catalytic cracking to gaseous products and isomerization to iso-octenes, the yield of the latter being 58.2 and 60.5 per cent by weight, correspondingly. To determine the constituents of the iso-octene fraction (b. p. 95—121.5°) obtained at 375° the latter was hydrogenated in the presence of a nickel catalyst, and the saturated product distilled on a column. From the distillation curve and refractive indices of several of the

* We have found that phosphoric acid on activated charcoal at 350—400° brings about the isomerization of cyclohexene into methylcyclopentene as well (unpublished data).

fractions the authors established the presence in the hydrogenated product of the following hydrocarbons: 2, 3, 4-trimethylpentane V (b. p. 113.8°), 2-methylheptane I (b. p. 117.2°), 4-methylheptane III (b. p. 118.0°) or 3, 4-dimethylhexane IV (b. p. 117.9°), and 3-methylheptane II (b. p. 119.0°):



Thus, according to Egloff *et al*. *n*-octenes under the action of the silica-alumina mass undergo at 375°C an extensive isomerization to methylheptenes, dimethylhexenes and trimethylpentenes.

We have studied the isomerization of octene-1 at 450° in the presence of beryllium oxide, and at 400° in the presence of silica gel.

Octene-1, prepared by the dehydration of octanol-1 over silica gel at 325° after purification by distillation through a column had the following properties: b. p. 121—122°/742 mm; $d_4^{20}=0.7166$; $n_D^{20}=1.4110$.

Its hydrogenation in the presence of platinized charcoal at 150° yielded *n*-octane with the following properties: b. p. 124—124.5°/749 mm; $d_4^{20}=0.7020$, $n_D^{20}=1.3984$.

Beryllium Oxide and Octene-1. 46.8 g of octene-1 were passed over beryllium oxide at 450° at a rate of 5 g per hour. The inner diameter of the reaction tube was 15 mm; the length of the beryllium oxide layer was 25 cm. The condensate obtained (43.4 g) was subjected to fractional distillation:

1. 35—113°	2.6 g	$n_D^{20}=1.4050$
2. 113—120°	37.8 »	» =1.4162

The main fraction 113—120° was again brought into contact with beryllium oxide under the same conditions. The condensate (35.4 g) yielded the following fractions:

1. 36—112°	2.4 g	$n_D^{20}=1.4112$
2. 112—120°	31.3 »	» =1.4174

Fraction 112—120° was again passed over the contact under similar conditions. The condensate (29.5 g) was distilled:

1. 36—112°	1.4 g	$n_D^{20}=1.4042$
2. 112—120°	26.0 »	» =1.4175

The products obtained were subjected to fractional distillation over metallic sodium (753 mm)

1. 34—95°	3.4 g	$d_4^{20}=0.6778$	$n_D^{20}=1.3971$
2. 95—113°	1.3 »	» =0.7194	» =1.4150
3. 113—118°	18.6 »	» =0.7223	» =1.4161
4. 118—122°	8.5 »	» =0.7309	» =1.4208

It follows from these data that under the action of beryllium oxide octene-1 underwent considerable isomerization to hydrocarbons of a branched structure. Along with isomerization the formation of small amounts of low boiling decomposition products and high boiling condensation products had taken place.

From our previous investigations, as well as from physical constants of the main fractions of the isomerizate 113—118° and 118—122°, the products of the isomerization of octene-1 appeared to consist mainly of isomeric methylheptenes.

As the products of isomerization of octene-1 are difficult to identify either by chemical methods or by the method of fractional distillation, the individual composition of the isomerizate was investigated by the combination scattering spectral method. As there are no systematic data on the spectra of combination scattering of isomeric octenes, the products of the isomerization of octene-1 were first converted into isomeric octanes by means of catalytic hydrogenation over platinized charcoal at 150°. The spectra of combination scattering of the latter have been the subject of a systematic and detailed study ([7],[8]).

Isomeric octenes could be expected to be present in fractions 95—113°, 113—118° and 118—122°. Owing to the small amount of the first fraction, only the two last ones were investigated.

Hydrogenation of fraction 113—118° yielded a saturated product having the properties: b. p. 114—119°/748 mm; $d_4^{20}=0.7072$; $n_D^{20}=1.4000$. Found MR_D 39.15; calculated for C_8H_{18} 39.14.

Fraction 118—122° after complete hydrogenation had the properties: b. p. 119—122°/752 mm; $d_4^{20}=0.7099$; $n_D^{20}=1.4019$. Found MR_D 39.17; calculated for C_8H_{18} 39.14.

Fraction 114—119° when investigated by the combination scattering spectral method was found to consist mainly of 4-methylheptane (b. p. 118.0°) and 3-methylheptane (b. p. 119.0°); it also contained 2-methylheptane (b. p. 117.2°).

The combination scattering spectra of fraction 119—122° showed that it consisted of 3-methylheptane (b. p. 119°) and of *n*-octane (b. p. 125.6°).

Identification was made from the following data on the combination scattering spectra of isomeric octanes (only the main frequencies up to 1040 cm^{-1} having intensities 1 and higher are referred to).

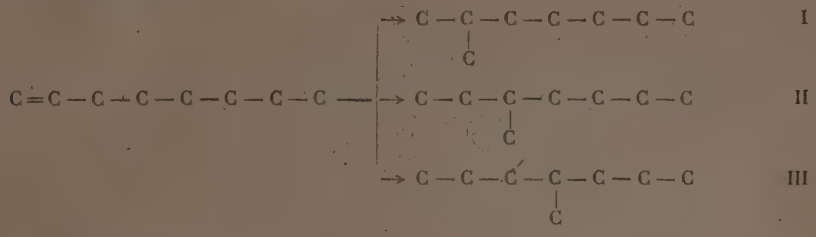
2-methylheptane (b. p. 117.2°)		4-methylheptane (b. p. 118.0°)		3-methylheptane (b. p. 119.0°)		*n*-octane (b. p. 125.6°)	
$\Delta\nu$ cm^{-1}	*I*	$\Delta\nu$ cm^{-1}	*I*	$\Delta\nu$ cm^{-1}	*I*	$\Delta\nu$ cm^{-1}	*I*
288	5	316	8	298	3	282	6
318	1,b	420	2	315	2,b	300	1,b
380	1,b	453	1	418	1,b	766	1
409	2	800	1	454	1	813	2,b
427	1	821	3	762	2	842	2
763	2	838	2	821	5	860	2
816	6	873	3	874	3;s	877	3
879	2	910	4,d	895	4,s	895	4
895	3	1010	2	908	3	954	1;b
952	4,s	1042	6	961	1,b	969	2
963	4,s			978	2	1028	1
				1040	3,b	1044	1

Notes. b, broad line; s, sharp line; d, double line.

The investigation was carried out on a three-prism Zeiss spectrograph with the middle camera having $f=27$ cm; dispersion of the apparatus for $\lambda=4390$ Å was 27 Å/mm; for $\lambda=4940$ Å, 47 Å/mm. Comparation was effected using an Abbe comparator (Zeiss), guaranteeing an accuracy of 0.001 mm. Data on intensities are expressed in an arbitrary scale, the intensity of the cyclohexane line 802 cm^{-1} being put at 60.

From the study of the combination scattering spectra of the hydrogenated products of the isomerization of octene-1 it follows that the main isomerization pro-

ducts are 4-methylheptene, 3-methylheptene, and 2-methylheptene. Hydrogenation of these olefins lead to the formation of 4-methylheptane III,3-methylheptane II, and 2-methylheptane I.



Thus, optical analysis of the isomerisate confirmed our conjecture that when octene-1 was brought into contact with beryllium oxide, the migration of the double bond took place with a simultaneous isomerization of non-forked octenes into methylheptenes.

Silica Gel and Octene-1. 14.3 g of octene-1 were passed three times over silica gel at 400° at a rate of 4 g per hour. The inner diameter of the reaction tube was 15 mm; the length of the silica gel layer was 50 cm. The condensate obtained (9.3 g) was subjected to fractional distillation over metallic sodium (740 mm).

1. 35—113°	0.8 g	$n_D^{20}=1.3990$	
2. 113—118°	6.5 »	» $=1.4169$	$d_4^{20}=0.7246$

The main fraction 113—118° was hydrogenated over platinized charcoal at 150°. The hydrogenated product had the properties: b. p. 113—117°/752 mm; $n_D^{20}=1.3994$; $d_4^{20}=0.7062$. Found MR_D 39.15; calculated for C_8H_{18} 39.14.

As may be seen from these data, silica gel at 400°C also brings about the isomerization of octene-1 into hydrocarbons of a branched structure. The physical properties of the iso-octene fraction 113—118° as well as of its hydrogenation product are very similar with those of the corresponding products obtained in experiments with beryllium oxide. The conclusion may be drawn that the product of the isomerization of octene-1 in the presence of silica gel consists also mainly of a mixture of isomeric methylheptenes.

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GEOCHEMISTRY

ON THE ORIGIN OF THE URAL BAUXITES

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1. Concerning the origin of the Ural bauxite ores the opinion generally accepted now is that of A. D. Arkhanguelsky (1). According to the above-named author, these ores are aquatic sediments, the Palaeozoic bauxites having been deposited in the sea, and the Mesozoic ones in lakes and swamps. Lateritic weathering is said to have been the source of alumina, iron oxides, TiO_2, and silica brought, in form of solutions, either ionic or colloidal, from the continent to the sea, lakes or swamps.

To prove if the above-mentioned considerations are valid, let us take as an example the Red Cap bauxite deposits situated in the Urals about 60° N. The Red Cap bauxites, Middle Devonian in age, are believed by Arkhanguelsky to be marine sediments.

However, all we know about the Red Cap bed is not consistent with such an opinion. The bauxites of this place were deposited on a plain formed by marine Devonian limestones and dotted with numerous karst depressions. The bauxite bed is represented by a series of isolated blocks, numbering seven over a distance of six km.

Spaces between blocks devoid of ore and measuring 0.1—4 km represent elevations of the primary surface where no bauxites were deposited. The bauxites have been accumulating only in depressions of the Devonian plain (2).

Further at the base of some Palaeozoic bauxite beds in the Urals (Red Cap, Ivdel, Bogoslovsk) there is present a peculiar bauxite breccia. In the Red Cap bed it consists of angular limestone fragments measuring 5—10 cm and cemented together by a micro-oolitic bauxite substance. This breccia is, without any doubt, result of subaerial weathering of limestones, or their eluvium, which was afterwards cemented by the bauxite substance carried away from the lake or swamp deposits.

It will be seen from the aforesaid that the Red Cap bauxites were deposited not in the sea but on the continent. It is clear that a terrain composed of limestones was dotted with numerous lakes and swamps situated in depressions of an old karst landscape, and serving as containers for the deposition of bauxite sediments.

The occurrence of corals in the Ivdel chamosite-diaspore ores is the chief argument for the opinion that some Ural bauxites are of marine origin. Let us examine the exposure on the board of the Ivdel River (about $60^1/_2°$ N lat.) where corals are said to be found in the bauxite ore. According to Marcova and Streis(3), we have here in descending order:

3. Bauxites without fauna.

2. Bauxite breccia: «a rock consisting of angular fragments of limestone cemented by bauxite substance. Many corals belonging to the genus *Favosites*».

1. Underlying limestones.

It is obvious from this description that corals were found not in bauxites, but in limestone fragments forming the bauxite breccia, *i. e.* in the eluvium of underlying limestones.

Corals are common indeed in limestones underlying the Devonian bauxites of the Urals.

Of the same type are obviously founds of the marine fauna in other bauxite sites, for instance, in the vicinity of the Red Cap (4). The composition of the marine fauna collected here does not permit to postulate its occurrence in a sea where a mass deposition of aluminium combinations took place. The latter are known to be detrimental for organisms. This fauna is doubtless synchronous with the deposition of limestones rather than with that of bauxites which fill up the karst depressions among the limestones.

According to Arkhanguelsky, the lateritic weathering supplied to the sea solutions of aluminium combinations. We know that in the process of laterite formation alumina is accumulating on the continent, whereas silica is carried away to the rivers and finally to the sea. Now the question arises, Where did the silica disappear. If silica is carried away together with alumina, it must give origin to clay and not to bauxite. It is incomprehensible why so large quantities just of alumina solutions were carried away from the weathering crust, alumina being practically insoluble in the water. «The short field of an independent existence of Al_2O_3 and of its solubility in conditions both of acid and alkaline media», Ginsburg says (5), «limits all possibilities for the solution and migration of alumina».

Thus, I am of the opinion that the Ural Palaeozoic bauxites were deposited on the continent, in a lake-swamp environment, like the Mesozoic Ural bauxites which, according to the current opinion, were formed in the same manner.

II. I suppose that the accumulation of alumina in the Ural bauxites has been accomplished by the swamp or lake-swamp vegetation. Let us substantiate this view.

Some plants, especially such as are growing in lakes and swamps, possess the capacity for accumulating alumina. Among such plants are the *Lycopodinae* comprising, besides the recent ones, the Carboniferous coal-producers, lepidodendrons and sigillarias, very characteristic for the Middle Carboniferous. They were trees inhabiting swamps. The first *Lycopodinae* and at the same time the first tree-like plants have recently been described from the Upper Silurian of Australia. The ash of some durain coals contains about 50 per cent of Al_2O_3. The ash of modern *Lycopodiales* contains also much alumina, not seldom more than 1 per cent to weight (6). The ash of a recent *Lycopodium* contains 52 per cent of Al_2O_3 and 13.6 per cent of SiO_2 (Lessing, 1926).

Certain ferns contain much alumina, especially in their underground parts; for example, *Aspidium filix-mas* contains nearly 10 per cent.

We may point out the following remarkable thing. In the Lower Carboniferous bauxites of Tikhvin (Leningrad region) there occur stigmarias—root organs of sigillarias and lepidodendrons. The lower horizons of the bed contain carbonized stigmarias, whereas the stigmarias in the upper horizons do not differ from bauxites either in respect of shape or chemical composition (7). It is obvious that the same vegetation under certain conditions—which we do not know in more detail—was able to give origin to coals, while in others it gave rise to bauxites. The Mesozoic Ural bauxite beds often include coal, and remains of *Equisetales* occur also.

We may lastly remind that there exist bacteria accumulating alumina; the ash of *Crenothrix* contains up to 50 per cent of alumina (8).

The accumulation of alumina may be observed in superficial horizons of recent podzolized bog soils. According to Rode (9), the mentioned type of soil developed on banded clays in the Leningrad region contains at a depth of 30 cm 30 per cent of Al_2O_3, calculated for waterless and humusless matter, whereas the amount of alumina in the underlying banded clays does not exceed 18—20 per cent.

Bauxite contains much titanium dioxide; in some specimens from Siberia the amount of TiO_2 surpasses 10 per cent. There are reasons to suppose that such an

accumulation owes its origin to biochemical processes. Titanium was found in the ash of both *Lycopodiales* and many terrestrial and aquatic plants ([10]). Rode assumes the probability of biochemical migration of titanium in podzolized soils.

Up to 8 per cent TiO_2 are accumulated in the upper horizons of the weathering crust of Chelyabinsk granites, Siberia.

Conclusions. The bauxites represent swamp (or respectively bog) or lake-swamp (bog) deposits. They were formed through accumulation of alumina and TiO_2 by swamp (bog) or lake vegetation, possibly under participation of bacteria.

Alumina was extracted by plants mainly from bottom deposits *in situ*. There is no need to admit a transportation of products of lateritic weathering from the continent down to swamps and lakes.

It is hardly possible to support the opinion that bauxites may accumulate among calcareous muds of the open sea.

Searches for bauxite ores are to be carried out among continental deposits. Localities where old karst landscape is developed are especially favourable.

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GEOLOGY

ON THE AGE OF THE ORE-BEARING STRATA OF THE ULU-TELYAK MANGANESE DEPOSIT OF BASHKIRIA

(FROM DATA OF MINERALOGICAL ANALYSIS)

By S. G. SARKISSIAN

(Communicated by W. A. Obruchev, Member of the Academy, 29. V. 1944)

The Ulu-Telyak manganese deposit is located 14 km from the Asha metallurgical works. According to unpublished data by A. V. Khabakov, the geological structure of the area of development of the ore-bearing belt is as follows.

The oldest deposits are «reefogenic hydroactinoid limestones of the middle of the pseudo-*Schwagerina* horizon of the upper Carboniferous», exposed on the Asha River at the Asha works. These rocks show a slight manganese mineralization. The Lower Artinskian deposits are represented by limestones and marls and compose the Zmeinye Mountains. They contain spots and beds of manganese ores. The Upper Artinskian deposits are represented by marls and marlaceous limestones. They are but partly manganiferous in their central portions.

The Kungurian deposits are represented by gypsum and marlaceous clays and are most extensively developed in the northern portion of the deposit.

They are overlain by «a horizon of coarsely oolitic marlaceous limestones and brown-yellow marlaceous clays», among which ore-bearing strata have been discovered which are the primary source of the manganese mineralization* at Ulu-Telyak.

According to Khabakov, this horizon rests in its central part directly on Upper Artinskian limestones. In the Kungurian gypsum (Zmeinye Mountains) it overlies at places Lower Artinskian deposits. Thus, the ore-bearing strata are transgressively overlying different horizons of the Upper Carboniferous, the Artinskian and Kungurian stages. The mineralogical composition of Ulu-Telyak ore has been described in detail in a paper by Beliankin and Lapin (1).

The present author made as early as in March 1943 mineralogical analyses of the terrigenous component of specimens from a number of workings in order to establish the age of the ore-bearing strata. Along with this, the mineralogical composition of the Ufa and Kungurian deposits of the Ulu-Telyak region was determined with the object of correlating them with specimens containing manganese. The manganiferous rocks analysed are represented by limestones, marls and marlaceous clays.

The composition of the heavy minerals is as follows: ore minerals (mainly pyrite), almandine, zircon, tourmaline, rutile, picotite, titanite, chlorite, common hornblende, diopside, biotite. Of the light minerals, quartz and decomposed feldspars have been discovered.

* In addition to the type of mineralization described, Khabakov distinguishes two more generations: eluvial oxidized ores formed «in pre-Neogene times of the ancient weathering», and friable redeposited ores lying at the base of Tertiary clays.

The Ufa red rocks (collection of B. M. Keller and D. A. Vital from the basin of the Lemeza and Ardyghel) are characterized by the following heavy minerals: ore minerals (chiefly pyrite), almandine, zircon, picotite, tourmaline, chlorite, biotite, common hornblende, diopside, titanite, rutile, epidote and rarely zoisite.

The light minerals are represented by quartz, decomposed feldspars, albite (No. 8) from the angle of extinction in the symmetrical zone. The Kungurian rocks are poor in terrigenous minerals and contain almandine, zircon and rarely rutile. The light fraction is usually surcharged with sulphates.

From a comparison of the mineralogical composition of the deposits investigated one is brought to conclude that the manganiferous deposits resemble the Ufa deposits. Knowing the stratigraphy of the Ufa deposits of the Near-Ural area of Bashkiria, the Ufa deposits of the Ulu-Telyak region, and hence also the ore-bearing strata of Ulu-Telyak, may be classed as the lower members of the Ufa series.

It should be noted that some beds of clays have been found to be «in the nearest vicinity of the Ulu-Telyak deposit... a higher manganese content... has been found in some beds of clays in the lower series of the Ufa red-coloured formation» (after D. E. Aisenverg, quoted according to Khabakov, unpublished).

In addition, in the ore strata Khabakov discovered remains of a land flora and semi-fresh water ganoid fishes, which, according to this author, points to an Ufa «age» of the manganiferous deposits.

Thus, all the data available are evidence in favour of the ore-bearing manganiferous rocks of the Ulu-Telyak deposit belonging to the lower horizons of the Ufa series.

Owing to a scarce supply of fresh water, and hence of terrigenous material, the conditions of sedimentation of the lower horizons of the Ufa series in some regions of Bashkiria were not essentially different from the lagoonal regime of Kungurian basins; therefore Khabakov is quite right in attributing a lacustrine-lagoonal origin to the manganiferous rocks.

The lower horizons of the Ufa series are the most extensively developed in North Bashkiria; therefore search and prospecting work should be conducted in the first place in this region.

In the central (Ishimbay-Sterlitamak) and southern regions of Bashkiria the lower horizons are also developed alongside with the younger ones; hence the possibility is not excluded that manganese deposits may be discovered here, too.

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PLANT PHYSIOLOGY

ON THE INFLUENCE OF THE STOCK UPON THE SCION IN COTTON

By N. N. KONSTANTINOV

(Communicated by A. A. Richter, Member of the Academy, 1. IV. 1944)

Intraspecific, interspecific and even intergeneric graftings have been known long since in agricultural practice. Intraspecific graftings and, in many instances, interspecific ones, too, are being successfully done nearly in all cultivated plant species. Darwin (1) was the first to point out a certain analogy existing between crossing and grafting. I. W. Michurin and his followers have demonstrated the great rôle which may be possibly played by grafting in plant breeding (vegetative hybridization, the method of mentor, *etc.*).

In spite of the fact that grafting has been applied since a long time, the problem of interaction between stock and scion cannot be regarded as satisfactorily solved.

The components of the graft do undoubtedly deeply affect each other. Yakovlev (5) has shown in his experiments that even the pigments of the leaves may migrate in perennial fruit-trees from the stock to the scion, and *vice versa*.

Unsatisfactory results of experiments on grafting are often accounted for by a lack of due control of the development of the stock and the scion. Michurin who considered the modification of the properties of the scion as a result of the influence of the whole foliage and the root system of the stock, repeatedly emphasized the importance of the activity of the leaves of the stock. Michurin writes as follows: «in the vegetative hybridization the main part is played by the work of the foliage system of the stock for a very young hybrid cutting or eye of another kind of plant» (2).

Čajlachjan (4) and Moshkov (3) have shown that the regulation of the amount of foliage of the stock and scion has a decisive effect when grafting is intended to hasten fructification.

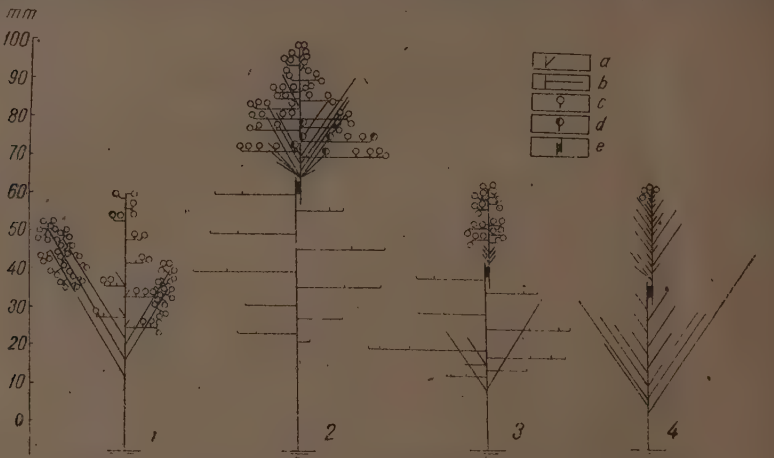
An attempt to control the development of cotton by means of grafting has given excellent results. It has been shown that with the aid of grafting blooming may be induced in a cotton variety which does not bloom naturally under a long-day regime in Tashkent; blooming may also be hastened in a late variety. It is necessary for this purpose to graft such varieties to an early one. Cuttings should be taken from very young plants, while the stock should be at the period of reproduction. After grafting the scions should be forced to development at the cost of assimilates produced by the stock. This is attained in practice by removing all the leaves of the scion at an earliest possible age. On the other hand, all the flower buds should be removed from the stock, and branch-ends broken, in order to restrain somewhat the growth of the stock and enhance the flow of food into the scion.

Such relations existing between stock and scion do not only hasten the development of the scion so that varieties which do not flower naturally are brought to blossom, but the scion acquires certain characteristics of the stock.

In this connexion it is of interest to report on the experiment with the varieties of cotton of the late type («Mid», *etc.*) which have been carried on by us.

The varieties tested were grafted at the cotyledon age by side grafting method below the cotyledons. The stocks used were as follows: 1) arborescent cotton which is incapable of blooming naturally under conditions of the long day at Tashkent; 2) ultra-early cotton of the annual type; 3) F_1 hybrids from the cross *Gossypium hirsutum* L. × *G. barbadense* L. All the stocks, except for the arborescent cotton, had already attained the reproductive phase towards the moment of grafting. As a control served plants grown under ordinary field conditions and of the same age as the scion*. After the scions and stocks had grown together, the development of the experimental plants was controlled in the manner indicated above, *i. e.* all the leaves of the scions were removed at an earliest possible age, while all the flowering buds were removed from the stocks.

Since all the plants were grafted beneath the cotyledon nodes, the formation of their aboveground parts was not disturbed by grafting. As may be seen from the figure, the height of formation of the first fructiferous branch (h_s) of the scion



a, monopodium; *b*, sympodium; *c*, flower bud; *d*, flower; *e*, place of grafting

	1. Control	2. Stock: F_1 *Gossypium hirsutum* × ×*G. barbadense*	3. Stock: *G. hirsutum* early	4. Stock: *G. hirsutum* arboreous
Formation of flower buds	24.VII	20.VII	18.VII	12.VIII
Blossoming	—	17.VIII	22.VIII	—
Number of flower buds	63	131	21	4
Number of flowers	—	4	1	—
h_s	11	6	5	15

depends on the stock upon which the plant developed. The first fructiferous branch has developed in the control from the 11th node. Under the influence of the early stocks there was observed a lowering of the first fructiferous branch down to the 5th or 6th node of the main stem (plants 2 and 3). As a result of the formation of the first fructiferous branch at a lower level, these scions produced earlier flower buds and bloomed earlier. Despite a certain delay in development caused by grafting, the scion grafted to the ultra-early variety (plant 3) attained the reproductive phase 6 days earlier than the control, and bloomed on 22.VIII, while the control plant was unable to bloom until the first frost set in (19.IX).

* For grafting little plants were used which were removed in thinning-out the control.

When cotton of the arboreous type was taken as stock, there was observed an extremely great delay in the development of the scion. In this case (plant 4) the first fructiferous branch was formed four nodes higher than in the control, and the scion attained reproductive stage 20 days later. It should be noted that the scion began reproducing in this experiment only owing to imperfect experimental technique. If the assimilatory activity of the scion could be absolutely eliminated (assimilation by the young leaves before their removal, green organs of the flower buds and flowers, green parts of the stem), the plant would have completely lost its assimilatory ability, and could have developed according to the type of the arboreous stock.

The change was found not to be limited to the rate of development. The study of our experimental plants has shown that the reproductive power of scion was depending on the stock. Under the influence of the stock, a F_1 hybrid characterized by high yielding capacity, the scion produced twice as many reproductive organs as the control. The graft to an ultra-early stock (plant 3) distinguished by its rather weak growth and slight fructification also grew but slowly, and formed one third of the fructiferous organs developed by the control. The scion grafted to an arboreous stock had produced only 4 flower buds.

The general appearance of the plant is materially changed under the influence of the stock. As shown in the figure, the experimental variety of cotton is distinguished by the presence of 4—5 monopodial branches in the lower part of the plant and by shortened fructiferous branches. Under the influence of the arboreous stock the plant assumed the monopodial habitus typical of the arboreous stock.

The stock, F_1 hybrid distinguished by its robust development and a spread-out habit of growth, transmitted this characteristic to the scion. As may be seen from the figure, instead of the compact growth typical of the variety tested, the scion plant acquires a more spread-out habitus under the influence of the stock.

The results of our study of *Gossypium patens* Cook et Hubb. are of great interest. It is indicated in botanical descriptions of this species ([6], [7]) that under the conditions of Mexico (Sonora, Sinaloa, 25—30° northern latitude) where it is distributed, it forms sometimes double bracts; two rings of three-leaved bracts should thus be formed. Under the conditions of Tashkent (41° northern latitude) where the summer day is longer, all the bracts formed by *Gossypium patens* are doubled. Experiments have shown that when the duration of the day is cut down to 8—10 hours, the plant forms exclusively normal three-leaved bracts. It follows that this character of *Gossypium patens* which had been included by the authors who described this species into the specific diagnosis, is extremely labile and, as shown by experiments, intimately dependent on the length of illumination.

It would be erroneous, however, to ascribe the appearance of this character to some specific action of the photoperiod. In order to find out the nature of the character in question, we have made graftings of *Gossypium patens* at the cotyledon age to other species in such a way as to bring up the scions on the assimilates of the stock. In this case the scion has formed only normal three-leaved bracts under the conditions of the normal long day of Tashkent.

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PLANT PHYSIOLOGY

RESPONSE OF CERTAIN WILD-ROSE SPECIES TO HIGH SALT CONCENTRATIONS IN SOILS

By I. A. KRUPENIKOV

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In the forests of the southernmost part of the north-western Kazakhstan (Naurzum wood, Aman-Karagay, Sypsyn-Agach) the processes of gradual increase in the concentration of salts in the soil covering go on very intensely. This is because of a lack of drainage in the region, and of extensively developed phenomena of continental salt accumulation and atmospheric impulverization of salts. Certain arboreal plants and shrubs (pine, kirghizian birch, aspen, hawthorn, *etc.*) do practically thrive on such soils, in spite of the high concentration of salts in them. The adaptation of these species to soils high in salt had been worked out in the course of their physiological and ecological remodelling, as a result of which peculiar «salt-marsh» plant ecotypes were evidently developed.

Different species of a single genus which under usual conditions are not supposed to be salt-resistant, do not respond in a similar way to high concentrations of salts in the soil; in accordance, their salt-resistance varies within a wide range. Certain representatives of the genus *Rosa* go well to illustrate this observation. Three species of wild-rose; *Rosa acicularis* Lindl., *R. laxa* Retz., *R. glabrifolia* C. A. Mey (*R. pratorum* Sukacz.) occur in the insular forests referred to above.

We came across *R. acicularis* on practically saltless soils only, containing mere traces of Cl' or SO''_4. Unlike the two other species, *R. acicularis* is highly drought-resistant and thrives on arid arenaceous soils.

R. laxa prefers more humid habitats. This species does not infrequently occur in regions with salt-meadow soils, in association with such plants as *Betula kirghisorum*, *Inula salicina*, *Glycyrrhisa uralensis* and less often with *Elymus salsuginosus*. *R. laxa* goes on well and yields abundant fruit on soils whose upper horizons are rather high in Cl' (up to 0.324 per cent) and SO''_4 (up to 0.187 per cent) (see table, plot 1). In these soils the conventional concentration of NaCl in the soil solution from upper horizons attains 27.0 g per litre.

Such a concentration of NaCl, according to Vant-Hoff's law, makes the osmotic pressure of the soil solution rise as high as to about 20 atmospheres. As a matter of fact, it will obviously be higher still, for besides NaCl the soil contains some other soluble salts, too.

The lower soil horizons (down to the second half metre) are not so rich in salts. According to our observations, the root-system of *R. laxa*, similar to that of other wild-rose species, penetrates to a depth of one metre and even deeper. Being less drought-resistant than *R. acicularis*, the development of this species suffers a pronounced inhibition in the case of a high concentration of salts throughout the soil profile; nevertheless, it is not infrequently found to grow on primitive arenaceous soils, in depressions, amid feather-grass and couch-grass steppes.

Humidity and Chemical Composition of Water Extract from Soils Sampled at Habitats of Various Wild-Rose Species in the Naurzum Wood (in per cent to absolutely dry soil)

No. of plots	Description of plot and soil	Species and condition of plants	Depth of sampling in cm	Solid residue	Total alkalinity in HCO'_3	Cl'	SO_4''	Humidity	NaCl (convent. concentration in g per l)	Date
1	Peripheral part of salt-meadow	*Rosa laxa*, good condition, abundant fruit	0 — 10	1.14	0.10	0.324	0.187	20.0	27.0	11. VIII
			20 — 30	0.74	0.08	0.220	0.138	14.5	24.8	1941
	Salt-meadow soil		75 — 85	0.13	0.07	0.009	0.016	7.5	1.9	
			110 — 120	0.05	0.01	0.001	0.008	4.2	0.39	
2	Salt-meadow	*R. glabrifolia*, good condition, height 160 — 170 cm. Abundant fruit	0 — 10	3.20	0.17	0.425	0.841	22.6	30.9	26. VIII.
	Salt-meadow soil		20 — 30	1.80	0.12	0.357	0.524	17.1	34.3	1941
			60 — 70	0.85	0.09	0.057	0.267	10.0	9.3	
		R. laxa, inhibited condition, height up to 130 cm. Fruits dried unripe, leaves prematurely yellow	110 — 120	0.91	0.02	0.215	0.114	12.4	28.5	

R. glabrifolia is of frequent occurrence in regions with salt-meadow soils of various salt concentrations; occasionally it is met there along with *Elymus salsuginosus*, *Statice Gmelini*, *Aster trifolium* and others.

R. glabrifolia goes on well and yields abundant fruit on soils which, being rather high in salts, produce an inhibiting effect upon the development of *R. laxa* (see table, plot 2). The very fact of the occurrence of *R. laxa* in those regions should be accounted for by the high rate at which the salt-meadow increased in salts, for it seems evident that by the time *R. laxa* settled there, the soil must have been much lower in salts.

R. glabrifolia is more water-loving than the two other species of wild-rose.

It may thus be seen that among the three species of wild-rose analysed above, *R. glabrifolia* is distinguishable by its higher salt-resistance; it is also the most water-loving species under the conditions that prevail in the region here described. If considered from the ecological point of view, *R. glabrifolia* is mainly a meadow-rose, with a very wide geographical area of distribution. It is of interest to note here that *R. laxa* which belongs to the group of species of the Central Kazakhstan and Central Asia, has been found less salt-resistant.

Special ecotypes of *R. glabrifolia* and *R. laxa* seem thus to have developed on soils enriched with salts.

The salt-meadow type of *R. glabrifolia* should widely be introduced to regions with soils high in salts, insular forests and similar regions.

This species of wild-rose is of rather high industrial importance, first, as a vitamin-containing plant (up to 2.8 per cent of ascorbic acid in the pulp of the fruit, and as much as 0.3 per cent, in the leaves). Secondly, it is a valuable ornamental plant. And, thirdly, it constitutes the main feed of the black and white grouse, while the root of the plants serves for the manufacture of tea-substitute.

Conclusions. 1. According to their salt-resistance the wild-rose species analysed above can be arranged in the following series, in the descending order: *Rosa glabrifolia*, *R. laxa*, *R. acicularis*.

2. *R. glabrifolia* can grow on soils in the upper horizons of which the concentration of Cl' is not above 0.425 and that of SO''_4 not above 0.841 per cent. The respective figures for the lower horizons are 0.215 and 0.114—0.267 per cent. If recalculated to the conventional concentration, the content of NaCl in the soil solution is up to 30.9 g per litre, which corresponds to an osmotic pressure of more than 20 atmospheres.

3. *R. laxa* resists fairly well to rather high concentrations of chlorides (up to 0.324 per cent of Cl') in the upper soil horizons. The salt-resistance of *R. acicularis* is no high one.

4. The following regularity has been found to hold for the species investigated: the salt-resistance of a species is inversely proportional to its drought-resistance, *i. e.* xerophytic and halophytic tendencies are antagonistic in the evolution of species.

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PLANT PHYSIOLOGY

PHOTOSYNTHESIS AND GROWTH

By A. A. RICHTER, Member of the Academy, K. T. SUKHORUKOV and L. A. OSTAPENKO

In our previous communication (1) it has been shown for two subspecies of beet (*B. vulgaris* subsp. *esculenta* and s. *altissima*) that the photosynthesis of the leaf remains constant in spite of the changing rate of growth and accumulation of dry matter.

In this note we wish to present the results of our studies on photosynthesis in a number of plants; these studies were intended to verify in how far the conclusions on the independence of photosynthesis of growth reached in the case of beets should be extended to other plants.

Photosynthesis was measured after Warburg (2), the external conditions being

Table 1

Plant	Fresh weight of one individual plant, in g					
	Total weight	Root	Leaves	Petioles	Stems	Flowering heads
Decapitated	1325	682	266	132	845	—
Flowering	1312	130	163	69	544	406

Table 2

Date	Regime	Height in cm	Number of leaves	Photosynthesis mg CO_2	Respiration mg CO_2
29.VII	Decapitated	147	24	12.24	4.32
» »	Flowering	177	27	12.15	4.95
8.VIII	Decapitated	150	25	10.62	3.69
» »	Flowering	210	29	11.70	2.60
27.VII	Watering	187	24	10.30	2.39
» »	Control	130	21	8.50	2.52
28.VII	Watering	185	24	12.40	2.50
» »	Control	132	23	12.70	2.89
28.VII	Nitrogen	188	29	11.03	2.84
» »	Control	158	21	11.38	3.06
7.VIII	Watering	183	28	11.53	4.45
» »	Control	167	27	12.85	3.66
7.VIII	Nitrogen	210	30	12.74	5.11
» »	Control	167	27	12.85	3.66

the same as described in our previous communication. The way of growing the plants, and the external agents applied to influence them are reported below.

Helianthus annuus (sunflower variety No. 169 of Saratov) was grown in the experimental plot of this laboratory. Growth was enhanced by optimal watering and by nitrate supply, as well as by «decapitation», *i. e.* by the removal of the apex of the stem at the very beginning of the formation of the inflorescence.

In Tables 1 and 2 are given the numerical data.

Table 3

Plant	Date	Regime	Description of the plants		Photosynthesis mg CO_2	Respiration mg CO_2
			Height	Number of leaves		
Zea Mays	28. VII	Nitrogen	100	15	10.07	2.58
	» »	Control	77	9	9.72	2.05
Cucumis sativus	28. VII	Nitrogen	102	16	9.79	2.80
	» »	Control	85	15	10.39	2.86
	» »	Removal of flower-buds	75	16	12.05	3.40
	» »	Control	46	12	11.40	3.50
Zea Mays	7. VIII	Watering	183	28	10.10	3.31
	» »	Control	67	27	12.50	2.49
	» »	Nitrogen	210	30	10.89	4.10
	» »	Control	87	12	12.50	2.48
Nicotiana rustica	2. IX	Flowering	82	18	9.64	2.00
	» »	Removal of vegetation point	48	14	9.00	2.72
	» »	Repressed	24	11	9.95	2.28
Zea Mays	31. VIII	Two doses of nitrogen	135	13	7.37	1.48
	» »	One dose of nitrogen	68	13	7.02	0.71
Ricinus communis	31. VIII	Two doses of nitrogen	62	7	9.07	2.18
	» »	One dose of nitrogen	55	6	7.84	1.28
Rudbeckia sp.	9. VII	Long day	17	—	10.05	1.49
	» »	Short day	Rosette		10.20	1.59
	31. VIII	Long day	44	—	9.15	1.90
	» »	Short day	Rosette		9.80	2.21
Nicotiana tabacum	30. VIII	One dose of nitrogen	92	25	8.11	1.58
	» »	One-fourth dose of nitrogen	19	14	6.68	1.06
Soja hispida	30. VIII	One dose of nitrogen	14	10	8.00	1.52
	» »	One-fourth dose of nitrogen	7	8	9.55	1.70
Nicotiana tabacum	28. VIII	Long day	22	15	8.78	2.31
	» »	Short day	11	15	10.25	2.70
Phaseolus multiflorus	» »	Long day	203	23	7.70	1.17
	» »	Short day	18	14	7.27	1.46
Taraxacum kok-saghyz	6. IX	Tetraploid	—	—	15.09	—
	» »	Diploid	—	—	15.49	—

From Tables 1 and 2 it follows that there is no correlation between growth, weight and leaf number, and photosynthesis. In fact, along with great differences in growth the plants show identical intensity of photosynthesis, or nearly so.

Let us consider the results of «decapitation»: the decapitated plants have shown a considerable increase in weight of all their parts, especially of roots. Thus, the total weight of the plant is 146 per cent to that of the control, while the weight of

the root is 524 per cent of the control. Consequently, decapitation has caused material disturbances in the correlations between growth and distribution of materials, but had no influence whatever upon photosynthesis.

Indian Corn (*Zea Mays*), Cucumbers (*Cucumis sativus*). Makhorka (*Nicotiana rustica*). The plants were grown at the experimental plot of the laboratory. The following agents were applied to modify growth: watering, nitrogeneous fertilizers, removal of the growing points of stems and removal of flowers. In the case of *N. rustica*, moreover, plants were selected which showed inhibition of growth due to unknown causes. Indices of the condition of plants and data on photosynthesis are presented in Table 3. They make one arrive at a conclusion similar to that in the case of sunflower, *viz.* that photosynthesis is not correlated with the process of growth.

Rubber Plant Kok-saghyz (*Taraxacum kok-saghyz* Rodin). The plants were grown at the experimental plot of the laboratory; both tetraploid and diploid forms were tested. Tetraploids were distinguished by their faster growth and larger dimensions. The seeds were received from Prof. M. S. Navashin.

Data on photosynthesis are presented in Table 3. Here photosynthesis is also similar in diploids and tetraploids. Consequently, the internal factor, the chromosome number, which modifies the process of growth does not affect photosynthesis.

Indian Corn (*Zea Mays*), Castor Bean (*Ricinus communis*), *Rudbeckia* sp., Soja Bean (*Soja hispida*), Tobacco Plant (*Nicotiana tabacum*), Multiflora Bean (*Phaseolus multiflorus*). In the case of experiments on photoperiod reaction and on the influence of the various dosages of nitrogen upon development of the plants, the plants were grown in pots. It should be noted that with reduced dosages of nitrogen the leaves showed a markedly paler colour («nitrogen-chlorosis»).

The figures reported here indicate that in these instances, too, there was no interrelation between photosynthesis and growth.

On the basis of data derived with various plants placed under highly varied conditions we are justified in saying that the initial phases of photosynthesis, including carbon dioxide absorption and its reduction under evolution of oxygen, proceed independently of growth processes, or of the distribution of the organic matter within the plant.

In order to find out in how far photosynthesis is an autonomous process, further investigation is necessary.

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EVOLUTIONARY MORPHOLOGY

FURTHER OBSERVATIONS ON TRANSFORMATIONS OF OSTEOCLASTS *IN VITRO*

CULTIVATION OF BONE TISSUE OF RAT AFTER ADMINISTRATION OF PARATHOHORMONE

By A. RUMJANTZEW and L. BEREZKINA

(Communicated by I. I. Schmalhausen, Member of the Academy, 15. IV. 1944)

The action of the parathyroid hormone on the bones of experimental animals has been carefully examined in the recent time. It has been found that the histogenetic effect of the parathyroid hormone may as well be obtained in the bone tissue outside the organism. For this purpose we explanted fragments of the *ossis femoris* of a chick embryo (1) into a medium previously saturated with the parathyroid hormone. The results of such experiments have shown that both *in vitro* and *in vivo*, between the *spongiosa trabeculae* one may find an increased number of osteoclasts and the transformation of osteoblasts into spindle-shaped cells. Studying the differentiation of osteoclasts *in vitro*, in our preceding investigations (2) we came to the conclusion that the osteoclasts represent a specific differentiation of the osteogenic mesenchyme. They appear in the bone system only. This conclusion was based on the following observations. We did not succeed in attaining such experimental conditions as would permit us to obtain a pure strain of osteoclasts, or mainly its differentiation in pure cultures of osteogenic cells.

Glücksmann's (3) indication concerning the presence of osteoclasts in a bone developing *in vitro* under the influence of strain and pressure, requires further analysis. What he saw may possibly have been macrophages rather than osteoclasts. The former can often be found in bone tissue cultures, and also observed in the bone in the case of experimental *osteitis fibrosae*. They have recently been described by McLeen and Bloom (5); we have often seen this type of cells in the beautiful preparations of Ivanova (4). A detailed examination of the changes which osteoclasts undergo in tissue cultures, as well as the total of facts available, lead to the following conclusions: 1) the osteoclasts are unable to phagocytize, 2) they are not transformed into macrophages, 3) they are not divided but, having lived for a certain time in the form of almost motionless cells, they at last disintegrate and disappear. However, as far as there were rather few osteoclasts in the growth area of the explanted fragments, it seems to us quite indispensable to confirm our observations on such material as would permit to have a great number of osteoclasts in the growth area.

For this purpose we explanted fragments of tissue taken from leg bones of young rats previously administered for 7 to 14 days parathyroid hormone at doses sufficient for the appearance of *osteitis fibrosae*. Recent investigation carried on by the author and his collaborators has shown that in the lamellar bones of these animals, under the same conditions, one can see a considerable number of osteoclasts immediately under the proliferating epiphysal plate. After having carefully removed the epiphyse we took out the upper cellular layer with sharp scissors, transferred it into Ringer's solution, divided it into little fragments, and explan-

ted these according to Carrel's classical method. We used as a medium the blood plasma of a rat + Ringer solution + a small quantity of chick embryo extract. In all 50 cultures were prepared.

In each culture we found, after an interval of 24 hours, a well developed growth area consisting of various cellular elements.

During the first hours following the explantation actively moving amoeboid elements began to migrate into the medium; after 24 hours there appeared spindle-shaped cells and large slowly moving osteoclasts. In some of the cultures the growth of the mesenchyme elements started at once together with the appearance of osteoclasts. The first type, evidently, corresponded to the growth of a tissue taken from internal parts of the *os. femoris*, whereas the second originated from the tissue forming the area of disintegrating *substantia spongiosa* lying at the surface.

In a previous work we have already described all the cellular elements of the growth area; in the present paper we shall deal only with the behaviour of osteoclasts. For this aim it is advisable to examine fixed and stained preparations. A considerable number of osteoclasts was to be found in the growth area of our cultures. They are mostly rounded cells with an easily stained ellipsis-shaped nucleus which is a little smaller in size than the nucleus of stellated osteogenic elements forming the growth area. The nucleus is rich in chromatin and usually contains several granular chromatin particles. The protoplasm is always dense and homogeneous; it has no large vacuoles; it is dark coloured, and shows, as a rule, a tendency toward an oxyphylic staining reaction.

In all our cultures we usually find the large polynuclear osteclasts lying near the fragment, while the mononuclear osteoclasts of a rounded or pear-shaped form may migrate faraway from the fragments, and one may see them at the border of the growth area.

The nucleus of mononuclear osteoclasts is disposed excentrically and, therefore, it reminds of a large or, more exactly, gigantic hypertrophied monocyte. However, when observing them *in vivo*, a striking difference is visible. Whereas the monocytes and histocytes are mobile, the osteoclasts are almost motionless. It is especially characteristic of the multinuclear forms. The large multinuclear cells have always a rounded shape but, in some cases, when the osteoclasts expand on the slide, they stretch out processes, either straight or of a fantastic form, giving the cells the most variable shapes. However, the osteoclasts of rats did not show such unusual protoplasmic rizoides as have been described by us for the chick embryo osteoclasts. In some cultures in protoplasm of mononuclear osteoclasts we occasionally observed a hypertrophic centrosome, but such type of cells developed no membranes and had no active motion. We never happened to observe the division of osteoclasts in the growth area, neither could we notice their formation by way of fusion of separate cells. In some cases we found obvious traces of nuclear fragmentation in the osteoclasts in the growth area, but, on the other hand, such osteoclasts, as a rule, bore evident signs of degeneration. As in the case of chick osteoclasts, we failed to establish any osteoclast transformations, except for a certain change in form, neither into typical macrophages of tissue cultures, nor into mesenchymatous cells.

It could be expected that the mononuclear osteoclasts, being younger cells, had not lost their capacity to transform; but having investigated their behaviour *in vitro*, we had been led to conclude that they were extremely stable cells lacking the transformation capacity, *i. e.* cells strictly determined and having a limited potency of development. The fate of all osteoclasts *in vitro* is death. But they can survive for a rather long time. Thus, in some cultures we noticed well-preserved mononuclear osteoclasts in the growth area on the 5th—6th day of cultivation. Multinuclear osteoclasts could be observed in the cultures of second transfer, but they disappeared in the third. While investigafing the section of the cultures prepared, osteoclasts may be found among the cells of the osteogenic mesenchyme, but, even in the sections, we never happened to see their development from mesenchymatous cells. It seems that the osteoclasts, having once been formed in the system of osseous tissues, did not arise again under conditions of cultivation

in vitro. If this circumstance be compared with our data obtained with cultures from the *os. femoris* of a chicken, we become once more convinced that the leading factor of the osteoclast formation is the parathyroid hormone which is absent in the medium of our cultures. Consequently, our investigation of rat osteoclasts *in vitro* confirms our assumption as to the specifity of osteoclasts appearing in the bone system only, and representing a finite stage of osteogenic cells development. The comparison with hematogenous and histiocytic elements of the bone marrow, which were always present in our cultures, shows with special clearness the difference existing between those elements. The osteoclasts are always larger in size, their protoplasm is thicker, it is not basophylic, and they are unable to phagocytize; these facts have also been noted in the recent work of McLeen and Bloom (5).

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HELMINTHOLOGY

SEASONAL DESTROBILIZATION IN THE CESTODE *MESOCESTOIDES LINEATUS*, PARASITIC OF SILVER-BLACK FOX

By T. S. SKARBILOVICH

(Communicated by K. I. Skrjabin, Member of the Academy, 3 III. 1944)

In the course of his study of the dynamics of mesocestoidosis in silver-black foxes of the Moscow province, the present author came across a rather peculiar phenomenon concerning the biology of *Mesocestoides lineatus* (Goeze, 1782), parasitic in the intestine of silver-black foxes. This phenomenon consists in that during the autumn-winter period the parasite sheds its strobila, so that the intestine of the fox is found to contain only scolexes deprived of segments.

During the winter season we succeeded to discover in the excreta of foxes by the method of vital examination but a small number of mature segments. Thus, in January (1944) it was 4.72 per cent, in February (1943) 8.9, in March (1942) 2.9. From May to August (1942 and 1943) there was observed a rapid increase in the invasion with the mesocestoidosis, which attained 25—35.6 per cent. According to observations carried on in 1942, in September the extensiveness of the invasion came down to 12.6 per cent, in October it was 8.6 per cent, and in November no more than 1.14 and 1.8 per cent. A total of 366 animals were examined in December 1942, and in none of the cases segments of cestodes were found to be eliminated into the external medium.

Thus, if we were to content ourselves by the vital method of examination of silver-black foxes, it might safely be believed that the animals get spontaneously freed of their cestosis invasion. But this is not the case, as a matter of fact; the parasite has not left the host, but only shed the strobila.

The phenomenon here described has been confirmed by the method of *post-mortem* diagnosis. When dissecting silver-black foxes at the time of the seasonal sloughter which took place in November and December 1942 and 1943, strobilae of *Mesocestoides* were but very rarely discovered by us in the intestine of the animals, whereas scolexes were of frequent occurrence.

A total of 62 foxes were autopsied in November 1942 at the Pushkinsky Hunting Soviet Farm; four of them, *i. e.* 6.4 per cent, were found to carry isolated strobilae, while 9, *i. e.* 15 per cent, contained scolexes only.

28 foxes were dissected in 1943 at the same Soviet Farm; now strobilae were found in 6, *i. e.* in 21 per cent, whereas scolexes were discovered in the intestines of 15 foxes, *i. e.* in 54 per cent. The number of heads of the cestodes varied from 1 to 54, which makes an average of 16 heads per one fox.

At the Saltykovsky Soviet Hunting Farm such silver-black foxes were dissected and examined in 1942 after the seasonal sloughter as had been found invaded by *Mesocestoides lineatus* during the preliminary examination made in June and July. Out of the 16 animals autopsied five have been found to contain isolated strobilae, which makes 31.3 per cent, while scolexes have been found in 14, *i. e.*

in 87.5 per cent of the animals. The number of scolexes varied from 4 to 208, which comes to an average of 50 heads per one fox.

By dissecting silver-black foxes in November and December we thus were able to show that the cestodes remained in the intestine of the animals mostly in the form of scolexes alone, whereas strobilae occurred but rarely and in small numbers.

The phenomenon of shedding the strobilae during winter has been described by Fediushin (1) for the cestodes *Raillietina urogalli*, *R. cesticillus*, *Rhabdometra tomica*, parasitic in the intestine of grouses. He gave this process the name of «destrobilization». This writer considers it to be a special adaptation of the parasite to the organism of its hosts, the grouses; the latter have sharply to change their diet in winter, passing to less nutritive feed, consisting of coarse woody fodder, which has a bad effect upon their organism. The shedding by the parasite of a part of its organism at a period when the host is particularly short of feed prevents the exhaustion of the latter's organism, being at the same time beneficial for the survival of the parasite. And the loss of the generative part of the body at the winter season, when there are no intermediate hosts, does by no means affect the prosperity of the biohelminthic species.

The destrobilization we observe in *Mesocestoides lineatus*, parasitic of silver-black foxes, does not depend on the diet of these animals: being bred artificially, they are given the same kind of feed throughout the year. Unlike this, the phenomenon of destrobilization is of paramount importance for carnivorous mammals under natural conditions. By shedding strobilae during the winter season when for lack of intermediate hosts the chances for the development of the larval stages are as good as nil, the parasite succeeds in saving its generative elements. Besides, the scattering of mature strobilae during spring and summer only, when the surrounding medium abounds in intermediate hosts, promotes the propagation of the parasite, raising its chances in the struggle for existence.

One cannot help supposing that the process of destrobilization must also take place in number of cestodes whose development proceeds under the participation of intermediate hosts and which display seasonal development. It may well be assumed that in number of cases the reduced intensity or total lack of cestode invasions that is observed during the winter season is due to the seasonal destrobilization of the parasite rather than to its short life, as it is usually believed, or its total spontaneous elimination from the host's intestine.

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